

(ALGEBRA)

LECTURES DELIVERED TO POST-GRADUATE STUDENTS OF
CALCUTTA UNIVERSITY

BY

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PART III—CONTINUED FRACTIONS
PART IV—APPROXIMATE SOLUTION
PART V—MATRICES. RESULTANTS

Part's III-V



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PREFACE

The third and the fourth part of these lectures deal with some classical portions of Algebra. Obviously a strict selection has been necessary, and the author gave preference to those portions of Continued Fractions (Part III) which are connected with the theory of numbers. In Part IV (Approximation solution) much importance has been given to how to carry out the calculations. Of course this branch of Algebra has to be considered from quite a different point of view from that of General Algebra. By approximation numerical values should be found out in the quickest and easiest way; so the reader cannot get a real insight into the nature and importance of the different methods without knowing the difficulties a practical reckoner has to face. The hints given by the author for the simplification of calculations do not involve the use of slide rules, calculating machines and graphical methods as these expedients are not familiar to our students. In this as well in some other items a later edition of these lectures is expected to show some alteration.

Part V contains the most important theorems on matrices with application on Hermitian and quadratic forms. Here the student may have the satisfaction of seeing in a nutshell the greater part of what he learned on Analytic Geometry.

As in the prefaces of Part I and of Part II, I have much pleasure in delivering my thanks here to friends and kind collaborators. The Vice-Chancellor of our University, Syamaprasad Mookerjee, Esq., M.A., B.L., M.L.A., Barrister-at-Law, gave me every possible help to get these lectures printed in a very short time, and is fully entitled to the grateful thanks of the author as well of the students. Proofs and manuscripts have been carefully revised by Mr. A. C. Chowdhury, M.Sc., Research Scholar of Calcutta University. The Calcutta University Press kindly co-operated in carrying out the printing in the proposed time.

F. W. LAYI

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PART III
CONTINUED FRACTIONS



§ 1. GENERAL PROPERTIES OF CONTINUED FRACTIONS.

Let K be a field, S a subring of K , and A be a subset of K with the following property: If a class of residues $\neq (0)$ of S contains an element of A , this class contains also the inverse of an element of A . Hence if

$$\alpha, \alpha', \alpha'', \dots, \alpha_1, \alpha_2, \dots \quad (1, 1)$$

denote elements of A and

$$s, s', s'', \dots, s_1, s_2, \dots \quad (1, 2)$$

denote elements of S , then every element of A can be represented either by

$$\alpha = s + 1/\alpha' \quad (1, 3)$$

or by $\alpha = s. \quad (1, 3')$

If e. g., K is the field of the real numbers, S the ring of the integers, and A the set of the real numbers > 1 , the representation (1,3), (1,3') is always possible and s is uniquely defined by α as the greatest integer $\leq \alpha$.

But the conditions (1,3), (1,3') hold also for other sets A in fields K , so an investigation in general terms is helpful.

Let $\alpha_1 = s_1 + 1/\alpha_2 \quad (1, 4)$

$$\alpha_2 = s_2 + 1/\alpha_3$$

$$\dots \dots \dots$$

$$\alpha_n = s_n + 1/\alpha_{n+1}$$

then $\alpha_1 = s_1 + \frac{1}{s_2 + \frac{1}{s_3 + \dots \dots \dots s_n + \frac{1}{\alpha_{n+1}}}}$ (1, 5)

The representation of α_1 by (5) is said to be a *continued fraction*. The formulae (1,4) can be continued till an element α_n will be an element of S . If there is an m so that $\alpha_m = s_m$, then the continued fraction is *finite*, otherwise it is *infinite*. If α can be represented by a finite continued fraction,

it belongs to the quotient field Q of S , and every finite set of elements

$$s_1, \dots, s_n \quad (1.6)$$

of S defines an element of Q by the help of (1.5).

We now define other sequences of elements of S by the following formulas.

$$P_{-1} = 0, \quad P_0 = 1, \quad P_k = s_k P_{k-1} + P_{k-2}, \quad k = 1, 2, \dots \quad (1.7)$$

$$Q_{-1} = 1, \quad Q_0 = 0, \quad Q_k = s_k Q_{k-1} + Q_{k-2}, \quad (1.8)$$

$$D_k = \begin{vmatrix} P_k & P_{k-1} \\ Q_k & Q_{k-1} \end{vmatrix} \quad (1.9)$$

then from (1.7) (1.8) (1.9) it follows that

$$D_k = -D_{k-1}, \quad \text{and as } D_0 = 1, \quad D_k = (-1)^k. \quad (1.10)$$

From (1.9) and (1.10) it follows that P_k and Q_k have no other common factors in S than unities and that

$$P_k : Q_k = P_{k-1} : Q_{k-1} = (-1)^k : (Q_k Q_{k-1}). \quad (1.11)$$

Let a_1 be an arbitrary element $\neq 0$ of K , then we get a uniquely defined sequence of elements a_1, a_2, \dots, a_{n+2} by the equations

$$a_i = a_1 : a_{i+1} \quad (1.12)$$

From (1.4) we get by multiplying the equations with a_0, a_2, \dots respectively

$$a_i = a_1 a_{i+1} + a_{i+2}, \quad i = 1, \dots, n \quad (1.13)$$

From (1.4) (1.7) (1.8) we get

$$\begin{aligned} P_k a_{k+1} + P_{k-1} a_{k+2} &= P_{k-1} (a_1 a_{k+1} + a_{k+2}) + P_{k-2} a_{k+1} \\ &= P_{k-1} a_1 + P_{k-2} a_{k+1}. \end{aligned}$$

By the repeated application of this formula we see that for $i \leq k$

$$P_k a_{k+1} + P_{k-1} a_{k+2} = P_i a_{i+1} + P_{i-1} a_{i+2} = P_0 a_1 + P_{-1} a_2 = a_1, \quad (1.14)$$

and by making a similar calculation with the elements Q we get

$$Q_k a_{k+1} + Q_{k-1} a_{k+2} = Q_i a_{i+1} + Q_{i-1} a_{i+2} = Q_0 a_1 + Q_{-1} a_2 = a_2. \quad (1.15)$$

Hence

$$\begin{aligned} (P_k a_{k+1} + P_{k-1} a_{k+2}) : (Q_k a_{k+1} + Q_{k-1} a_{k+2}) \\ = (P_i a_{i+1} + P_{i-1} a_{i+2}) : (Q_i a_{i+1} + Q_{i-1} a_{i+2}) = a_1. \end{aligned} \quad (1.16)$$

If we multiply (1.13) with Q_{i-1} and (1.13') with $-P_{i-1}$ and add, then it follows from (1.10) that

$$(-1)^i a_{i+1} = a_i Q_{i-1} - a_{i-1} P_{i-1}. \quad (1.14)$$

From (1.12), (1.13'') and (1.10) it follows that

$$a_i - \frac{P_i}{Q_i} = \frac{(-1)^{i-1}}{a_{i+1}} - \frac{a_{i+1}}{Q_i}. \quad (1.15)$$

Hence if the continued fraction is finite, i.e. $a_{n+1} = 0$

$$a_1 = \frac{P_n}{Q_n}. \quad (1.15')$$

If (1.4) holds, the elements x_1, x_2, \dots, x_n are said to be the *elements* of the continued fraction; the quotients $P_i:Q_i$ are the *convergents* and a_{n+1} is called a *complete fraction*. As a_1 is uniquely defined by these elements, we shall denote a_1 , if a_{n+1} exists, by

$$a_1 = (x_1, \dots, x_n | a_{n+1}) = \frac{P_n a_{n+1} + P_{n-1}}{Q_n a_{n+1} + Q_{n-1}}, \quad (1.5)$$

and if x_n is the last element of the continued fraction,

$$a_1 = (x_1, \dots, x_n) = \frac{P_n}{Q_n}; \quad (1.5')$$

from (1.4), (1.5'), (1.5'') it follows that for $k \leq n$

$$a_k = (x_k, \dots, x_n | a_{n+1}), \text{ respectively} \quad (1.5'')$$

$$a_k = (x_k, \dots, x_n). \quad (1.5''')$$

Let P'_i and Q'_i be defined by

$$\begin{aligned} P'_{-1} &= 0, \quad P'_0 = 1, \quad P'_i = a_{i+1} P'_{i-1} + P'_{i-2}, \\ Q'_{-1} &= 1, \quad Q'_0 = 0, \quad Q'_i = a_{i+1} Q'_{i-1} + Q'_{i-2}, \end{aligned} \quad (1.16)$$

then the following formulae hold:

$$\begin{aligned} \begin{vmatrix} P'_i & P'_{i-1} \\ Q'_i & Q'_{i-1} \end{vmatrix} &= (-1)^i \\ a_i &= P'_i a_{i+1} + P'_{i-1} a_{i+2} \\ a_{i+1} &= Q'_i a_{i+1} + Q'_{i-1} a_{i+2} \\ (-1)^i a_{i+1} &= a_i Q'_{i-1} - a_{i+2} P'_{i-1} \end{aligned} \quad (1.17)$$

$$a_i = \frac{P_{i-1}}{Q_{i-1}} + \frac{P_i}{Q_{i+1}}$$

From

$$\begin{aligned} P_n &= a_n P_{n-1} + P_{n-2} & Q_n &= a_n Q_{n-1} + Q_{n-2} \\ P_{n-1} &= a_{n-1} P_{n-2} + P_{n-3} & Q_{n-1} &= a_{n-1} Q_{n-2} + Q_{n-3} \end{aligned}$$

$$\begin{aligned} P_1 &= a_1 & Q_2 &= a_2 \\ P_0 &= 1 & Q_1 &= 1 \end{aligned}$$

we get the representation of P_n, P_{n-1} and of Q_n, Q_{n-1} as finite continued fractions. Then

$$\frac{P_n}{P_{n-1}} = (a_n, \dots, a_2, a_1) \quad \frac{Q_n}{Q_{n-1}} = (a_n, \dots, a_2) \quad (1.17a)$$

- [1.2] There is a very close connection between the finite continued fractions and the algorithm of the h.c.f.

Let a_1 be represented by a finite continued fraction $a_1 = (a_2, \dots, a_n)$. Then $a_1 - a_2 = a_3 = a_n$. Hence $a_n \approx a_n - a_2 + 0$ therefore $a_{n-2} = 0$. From (1.17), (1.17a), (1.17b), (1.14) we get therefore

$$\begin{aligned} a_1 &= P_n a_{n-1}, \quad a_2 = Q_n a_{n-1} \\ (-1)^n a_{n-1} &= a_1 Q_{n-1} - a_2 P_{n-1}. \end{aligned} \quad (1.18)$$

Hence $a_1 \approx P_n / Q_n$ belongs to the quotientfield of S .

Every common factor of a_1 and a_2 is a factor of a_{n-1} and a_{n-1} is a common factor of a_1 and a_2 . Hence a_1 and a_2 have an h.c.f. and this can be represented linearly by a_1 and a_2 . Especially $a = P_n / Q_n$ is a representation by two relatively prime elements of S , as

$$P_n Q_{n-1} - Q_n P_{n-1} = (-1)^n.$$

Let every element of the quotientfield of S be representable by a finite continued fraction and let $a, a' \neq 0$ be two arbitrary elements of S . Then $a : a' = a$ can be represented by two relatively prime elements of S so that a can be represented in a linear and homogeneous manner by those elements.

Hence $a : a' = p : q$ and $pq' + qp' = 1$.

Therefore $a_1 = a_2 = 1$ and $a_1 + a_2 + \dots + a_{n-1} + a_n = 1$.

Hence $a' = 1/a = a_1 + a_2 + \dots + a_n = 1$. So the arbitrary elements a, a' of S have an $h \in I$ ($h' = a_1 h = 1$) and thus is represented in a regular and homogeneous manner by a' and a'' . From this consideration it follows that it is not possible to represent the quotients of the elements of an arbitrary π -ring by finite continued fractions, if in S the elements are facturable, but the factoring is not unique there in at least a quotient of two elements of S which can be represented by an infinite continued fraction only. As every element of S is represented by a finite continued fraction, the finite continued fractions do not form a field in these cases.

Let a function $N(a)$ which takes positive integral values on y be defined for every element $a \neq 0$ of S and to every pair of elements a, a' of S let there exist two other elements a_1 and a'' so that

$$a = a_1 a' + a'' \quad \text{and} \quad \text{that}$$

$$\text{either} \quad a'' = 0 \quad \text{or} \quad N(a'') < N(a'), \quad (1.19)$$

Then $a, a' = a_1 + a' a_2 = a_1 + a_2 + a' a_3 + \dots$ and as

$$N(a') > N(a'') > N(a''') \dots > 0$$

are all integral numbers, the sequence a', a'', a''', \dots must be finite. Hence a/a' is a finite continued fraction. From these considerations we get the following theorem.

Theorem. Let S be an π -ring containing 1 and let a positive integer $N(a)$ be defined satisfying the conditions (1.19) for every element a of S then the quotientfield of S will be formed by the finite continued fractions of S , and the highest common factor of two elements a_1 and a_2 of S is given by a_{n+1} in (1.18) where $P_n = 1/Q_n$ have significance given by (1.6), (1.7), (1.8).

The formulae (1.13) and (1.14) are instances of linear fractional substitutions with coefficients from S the determinant being ± 1 .

Let A and B be the matrices of substitutions of this kind

$$A = \begin{pmatrix} a_1 & a_2 \\ a_3 & a_4 \end{pmatrix} \quad B = \begin{pmatrix} b_1 & b_2 \\ b_3 & b_4 \end{pmatrix} \quad \det A = a = \pm 1 \quad \det B = b = \pm 1 \quad (1.20)$$

$$A^{-1} = \begin{pmatrix} a_4 & -a_2 \\ -a_3 & a_1 \end{pmatrix} \quad B^{-1} = \begin{pmatrix} b_4 & -b_2 \\ -b_3 & b_1 \end{pmatrix}$$

of a be transformed by A into a' , and a' be transformed by B into a'' , then it follows

(1) a'' is transformed by E in a , and $\det E = 1$

(2) a' is " " " A' in a , and $\det A' = 1$

(3) a'' is " " " AB in a , and $\det AB = \det A' = 1$

The product of two matrices has been defined in Part I, § 25, and it has been shown in § 28 that the determinant of a product of matrices is equal to the product of the determinants. This is the case which we have to consider here, an easily proved by direct substitution.

An element of K is said to be *equivalent* to a if we get it by transforming a by a linear fractional substitution with determinant ± 1 . From (1), (2) and (3) it follows that this equivalence satisfies the conditions of reflexivity, symmetry and transitivity (see Part I, § 13), and therefore this equivalence defines a partition of K into classes, so that two elements of K are equivalent if and only if they belong to the same class.

By $\begin{pmatrix} x+1 & -1 \\ 1 & 0 \end{pmatrix}$ the element 1 becomes transformed into x , hence all

elements of S are equivalent. From (1), (2), and (3) it follows that the elements x_1, x_2, \dots defined by (14) are all equivalent. So in the first case any every finite continued fraction x_1, \dots, x_n is equivalent to $x_n = x_1$, and therefore belongs to the class containing the elements of S .

It is a not necessary to make a distinction between proper and improper equivalence. In the first case $\det A = 1$, in the second case $\det A = -1$. As $\det A = \det A$ in I, 20, holds the notion of proper equivalence as well as the notion of improper equivalence is a symmetric one. By combining two equivalences of the same kind we get a proper equivalence, and by combining two equivalences of different kind we get an improper equivalence. Every element is properly equivalent to itself, for the matrix E has the determinant 1. If in a class of equivalent elements an element a is also improperly equivalent to itself, i.e. if a is transformed into a by E , and $\det E = -1$, then an arbitrary element b of the same class becomes transformed into a by BE and by BE . One of these matrices has the determinant 1, the other has the determinant -1 , so each element of the class is properly and improperly equivalent to a , and therefore every element a at the same time properly and improperly equivalent to every other element of the class. If on the other hand a becomes

transformed into l by A as well as by B where $\det A = 1$, $\det B = -1$, then a becomes transformed into $-a$ by BA , where $\det BA = -1$, and therefore a is improperly equivalent to $-a$ so that in this case every a has element of the class $-a$ properly and improperly equivalent to every other element of G . If in G there are n pair of elements properly as well as improperly equivalent then G must be divided in two classes without common elements: the elements of the 1st class are properly and the elements of the 2nd class are improperly equivalent to a . Elements of the same set class are properly equivalent, elements of different set classes are improperly equivalent. *

Illego

Theorem 1. In a class of equivalent elements, either every element is properly self-improperly equivalent to every other element or there are two sub-classes with out common elements so that elements of different sub-classes are improperly equivalent.

As Γ is transformed into itself by $\begin{pmatrix} 1 & 0 \\ 0 & 1 \end{pmatrix}$, every pair of elements α properly and improperly equivalent in the class containing the finite continued fractions

Let α be transformed into itself by A . Then

holds, hence $a_2 u^2 + (a_1 - a_2) u - a_3 = 0$. (1.21)

There are 8 different cases

If $\sigma_1 = \sigma_2 = \sigma_3 = \sigma_4 = 0$. In this case $A = 1$, or -1 .

By these transformations every element is transformed into itself and E generate proper equivalences.

2. $(1, 21)$ is a radical polynomial, so, this is possible only if ϵ is an element of the quotientfield Q of B .

3. $(1, 21)$ is reduced to $(1, 1)$. In this case $\alpha = 1$ is integral to Q and f is order 2 over Q .

It is these considerations it follows that not every effect is improperly equivalent to itself.

Let the elements a_1, a_2, \dots, a_n be defined by $a_i = 1 - i$, be not all different

$$2 = 2 \quad \text{check}$$

[illegible]

Hence α can be represented by an infinite periodic continued fraction with the period a_{i-1}, \dots, a_i . From (1.10) it follows that α^{-1} becomes

transformed into itself by the matrix $D = \begin{pmatrix} P_i & P_{i-1} \\ Q_i & Q_{i-1} \end{pmatrix}$ where

$\det D = (-1)^i$ and becomes therefore equivalent to itself by the transformation

(1.13'') it follows that $\alpha_i = \frac{P_{i-1} + P_{i-2}}{Q_{i-1} + Q_{i-2}}$ belongs to the field

$Q(\alpha)$. Hence α_1 satisfies an equation of degree 2 with coefficients from S . The same holds for $\alpha_2, \alpha_3, \dots$.

[1.4] Let now a periodic sequence $\alpha = (a_1, \dots, a_m, a_1, \dots, a_m, \dots)$

of elements of S be given. It is not certain that in any extension of the quotientfield Q of S there exists an element which can be represented by the infinite periodic continued fraction

$$(\alpha_1, \dots, \alpha_m, \alpha_1, \dots, \alpha_m, \dots), \quad (1.22)$$

If such an element exists, it is not certain that the element is uniquely defined in the field. But if there is a field in which there exists one and only one element α represented by the periodic continued fraction (1.22) then

$$\alpha = (\alpha_1, \dots, \alpha_m, \alpha_1, \dots, \alpha_m, \dots) = (\alpha_1, \dots, \alpha_m | \alpha)$$

holds and this is the case considered just before.

2. REPRESENTATION OF THE POSITIVE NUMBERS BY CONTINUED FRACTIONS

[2.1] Let the elements $1 \leq i \leq m$ of the set Λ be the real numbers ≥ 1

and let S be the ring of the integers. Then the representation by the form (1.3) and (1.3')

$$\alpha = \alpha + 1; \alpha' \quad \text{or} \quad \alpha = \alpha$$

is always possible. If α is not an integral number, there is

$$1 \leq \alpha < \alpha < \alpha + 1, \quad \alpha' = 1, \quad \alpha + 1 - \alpha$$

representation is unique and the continued fraction is infinite. If the number is rational, there exist two representations, one by an even number continued fraction and the other by an odd number continued fraction.

[2.2.] We will now prove that the converse theorem also holds: every sequence satisfying (1.1) and (1.2) has a unique real number.

Let a_1, a_2, \dots be an infinite sequence satisfying the conditions (1.1b). The numbers P_n and Q_n should be defined as follows: (1.7) and (1.8)

$$P_{-1} = 0, P_0 = 1, P_1 = a_1, P_2 = a_1 a_2 + 1; P_3 = a_1 a_2 a_3 + a_1 + a_2 + 1;$$

$$Q_{-1} = 1, Q_0 = 0, Q_1 = 1, Q_2 = a_2; Q_3 = a_2 Q_1 + Q_0$$

From $P_2 = a_1 a_2 + 1 > 0$ it is easily mathematically deduced that

$$Q_2, Q_3, \dots, P_3, P_4, \dots > 0 \quad \text{and that}$$

$$0 \leq P_1 < P_2 < P_3 < \dots$$

(2, 8)

$$Q_0 = 0 < Q_1 \leq Q_2 < Q_3 < \dots \text{ hold}$$

The quantities $\frac{P_k}{Q_k}$ for $k \leq n$ are the convergents of each other. The same is

$$(1.1) \quad \frac{P_n}{Q_n} = \frac{P_{n-1}}{Q_{n-1}} + \frac{a_n}{Q_{n-1} Q_n} \quad \text{We can therefore apply (1.1b) to } \frac{P_n}{Q_n} = \frac{P_{n-1}}{Q_{n-1}} + \frac{a_n}{Q_{n-1} Q_n}$$

Hence

$$\frac{P_k}{Q_k} - \frac{P_{k-1}}{Q_{k-1}} = \frac{(-1)^{k+1} a_{k-1}}{a_k Q_k} > 0 \quad \text{if } k \geq 1 \text{ odd}$$

$$< 0 \quad \text{if } k \geq 1 \text{ even}$$

Hence

$$\frac{P_1}{Q_1} < \frac{P_2}{Q_2} < \dots < \frac{P_{2m+1}}{Q_{2m+1}}$$

(2, 4)

$$\frac{1}{Q_2} > \frac{P_{2m}}{Q_{2m}} > \dots > \frac{P_{2m-1}}{Q_{2m-1}} \quad \text{for } m = 1, 2, 3, \dots$$

The quantities $\frac{P_n}{Q_n}$ form therefore two sequences, one is increasing, the other is decreasing and every number of the first sequence is less than every

number of the second one. The intervals $\left(\frac{P_n}{Q_{n+1}}, \frac{1}{Q_n} \right)$ form therefore a

set of intervals, each of them is bounded by two consecutive

$$\text{As } \frac{P_n}{Q_n} - \frac{P_{n-1}}{Q_{n-1}} = \frac{(-1)^n}{Q_n Q_{n-1}}, \quad \text{[see (1.11)]}$$

The length of these intervals converges to 0. Hence there is for every given sequence of 1's, an arbitrarily positive number ϵ such

$$\frac{P_n}{Q_n} - a < \frac{1}{Q_{n-1}} \quad \text{for } n > 1. \quad (2.5)$$

a a positive number. This is the number which is not determined by the sequence a_1, a_2, \dots and can't be represented by a continued fraction. We will prove that this continued fraction is equal to $a_1 + a_2 + \dots$. It is not obvious from the above that it was that if there exist a number which is represented as a continued fraction $a_1 + a_2 + \dots$ then this number can only be the number 1. In fact, that sequence is there and no other number

belonged within all the intervals $\left(\frac{P_n}{Q_n}, \frac{P_{n+1}}{Q_{n+1}} \right)$, but it may so happen that

there is no such number and that the representation of a as a continued fraction furnishes another sequence which defines the same real number a . Of course this duplication cannot occur. We will show this by the following lemma which gives some idea of the distribution of the continued fractions on the axis of the real numbers.

Lemma. Let P_n/Q_n be the convergents of (a_1, a_2, \dots) , and P'/Q' be the last convergent of $a_1 + a_2 + \dots + a_n + t$, where $n = 1, 2, \dots$ and $t > 0$. Then

$$P_n/Q_n < P_{n+1}/Q_{n+1} \leq P'/Q' \quad \text{if } n \text{ is odd,} \quad (2.6)$$

$$\text{and } P_n/Q_n > P_{n+1}/Q_{n+1} \geq P'/Q' \quad \text{if } n \text{ is even.}$$

$$\text{Proof.} \quad Q_{n+1} = a_{n+1} Q_n + Q_{n-1}$$

$$Q' = (a_n + t) Q_{n-1} + Q_{n-2} = Q_n + t Q_{n-1}$$

$$\frac{1}{t} Q' = \frac{1}{t} Q_n + Q_{n-1} \leq Q_{n-1}$$

The equality holds only if $n = 0$ or if $t = a_{n+1} = 1$.



If n is even, the notation \leq must be replaced by $<$. For the continued fractions having only one even term, then $s_1 = r_1 = 1$. So we get a best partition of the rationals by the theorem (1.2). The continued fractions $(s_1, \frac{1}{s_1})$ begin $(s_1 = 1), (s_1 = 2), (s_1 = 3), \dots$ and $r_1 = 1$.

For the continued fractions $(s_1 = 2), (s_1 = 3), (s_1 = 4), \dots$ then having two limiting points. The continued fractions $(2, s_2)$ are situated between $(1, 2) = 3 + \frac{1}{2}$ and $(1, 3) = 3 + \frac{1}{3}$. The continued fractions $(3, s_2)$ are situated between $(1, 3) = 3 + \frac{1}{3}$ and $(1, 4) = 4 + \frac{1}{4}$. The limiting point is the proper part. If we consider the intervals between s_1 and s_2 every segment (s_1, s_2) of the odd division is divided by the $(s_1 + 1)$ th division into $s_2 - s_1$ segments, $s_2 - s_1$ is an even number, so we get limiting point and then s_1 and s_2 are odd and even.

$$P_1 = 1, P_2 = 2, P_3 = 3, P_4 = 4, \dots, 1, Q_1 = 1, Q_2 = 2, Q_3 = 3, \dots$$

$$Q_1 = 1, Q_2 = 2, Q_3 = 2s_1 + 1$$

$$s_2 = 1, 2, 3, 4, \dots$$

$$\frac{1}{s_2(s_2+1)} = \frac{1}{1 \cdot 2}, \frac{1}{2 \cdot 3}, \frac{1}{3 \cdot 4}, \frac{1}{4 \cdot 5}, \dots$$

As every rational number is represented by a continued fraction, and especially by a continued fraction which has even terms with an odd number of terms s_1 with s_2 or $s_2 + 1$. For every s_1 and s_2 we get the limiting point of one odd and even and of one even and even. So we get a natural partition of the rational numbers. Numbers being represented by $P_1, Q_1, P_2, Q_2, P_3, Q_3, \dots$.

The importance of this construction is shown by the following theorem.

Theorem. If $s > 0$, and $\frac{P_{2n-1}}{Q_{2n-1}} < \frac{s}{s} < \frac{P_{2n+1}}{Q_{2n+1}}$ then

$$s > Q_{2n} > Q_{2n-1}.$$

Proof. From the supposition it follows directly that

$$0 < \frac{s}{s} - \frac{P_{2n-1}}{Q_{2n-1}} < \frac{P_{2n}}{Q_{2n}} - \frac{P_{2n-1}}{Q_{2n-1}} = \frac{1}{Q_{2n}Q_{2n-1}}$$

and as s and Q_{2n-1} are positive

$$0 < s - Q_{2n-1} < \frac{1}{Q_{2n-1}}$$

the middle part of this inequality is an integer positive number

$$\text{Hence } 1 \leq \frac{s}{Q_{2n-1}} \text{ i.e. } Q_{2n-1} \leq s$$

$$1 \leq \frac{s}{Q_{2n-1}} < \frac{s}{Q_{2n-1}} + \frac{1}{Q_{2n-1}} = \frac{s + 1}{Q_{2n-1}}$$

As these numbers tend to $\pm\infty$, a proximation of a real number by the set of fractions with limited positive denominators are the proximations by the sequence $1, 2, 3, \dots$

§ 3. PERIODIC CONTINUED FRACTIONS WITH INTEGRAL COEFFICIENTS.

3/1) Let α be a real number such that the periodic continued fraction in S represents one number α in S and not a \sqrt{D} or rational solution of a quadratic equation in S . In the general case where S is the ring of the integers in \mathbb{R} , it is shown that the sequences 2 represent not only one positive real number α but also $-\alpha$ that the periodic continued fraction (2) represents exactly quadratic to the field \mathbb{R} of the rational numbers. However, it is not so in \mathbb{R} .

Theorem. If a positive irrational number α belongs to a field \mathbb{R} , where $[\mathbb{R} : \mathbb{Q}] = 2$ and α is represented by a periodic continued fraction

Proof. α has to be the root of polynomial

$$ax^2 + 2bx + c; \quad (3.1)$$

and the other limit $-\alpha$ is represented by a continued fraction $u = (s_1, \dots, s_n, \lambda)$. From (1.13') it follows that

$$\frac{P_n(\lambda) + P_{n-1}}{Q_n(\lambda) + Q_{n-1}}$$

is equal to $P_n(\lambda) + P_{n-1}^2 + 2P_{n-1}\lambda + P_{n-2}^2 + Q_n(\lambda) + Q_{n-1}^2 + 2Q_{n-1}\lambda + Q_{n-2}^2 = 0$

α is a root of
$$p(x) = 2(hx + k)x + l, \quad (3.2)$$

where
$$\begin{vmatrix} j & h \\ h & k \end{vmatrix} = \frac{P_n - P_{n-1}}{Q_n - Q_{n-1}} = \frac{l}{b} = \frac{c}{b} \quad (3.3)$$

Let β be the second root of (3.2) and μ be defined by

$$\mu = \frac{P_n - P_{n-1}}{Q_n - Q_{n-1}} = \frac{P_n}{P_{n-1}}, \quad \text{hence} \quad (3.4)$$

$$\beta = \frac{P_n\mu + P_{n-1}}{Q_n\mu + Q_{n-1}}$$

As $(x^2 + 2l/x + c) = 0$, the nature of μ is clear. From (3.1) it follows that

$$\mu = \frac{Q_{n-1}}{Q_n} = \frac{1}{(Q_n - Q_{n-1}) - P_n}$$

$$\text{As } \frac{P_n}{Q_n} = \frac{P_{n-1}}{Q_{n-1}} + \frac{P_{n-1}}{Q_n} = \frac{1}{Q_{n-2}Q_{n-1}} + \frac{1}{Q_n^2} \\ = \frac{P_{n-1}}{Q_n} + \frac{1}{Q_n^2}, \text{ where } |s| < 1.$$

$$\text{So } P_n = Q_n s = \frac{1}{Q_n}.$$

From (3, 2) it follows therefore

$$\begin{aligned} p &= \frac{Q_n}{Q_n} + \frac{1}{Q_n^2} + \dots \\ &= \frac{1}{Q_n} + \left(\frac{1}{Q_{n-1}Q_n} + \frac{1}{Q_n^2} + \dots \right) \\ &< 0 \text{ for } Q_n, Q_{n-1} | B-n | > 2 \end{aligned} \quad (3.6)$$

Hence after a certain integer n the p, q become negative, therefore $k-j$ also becomes negative. As $h^2 - c = l + k$ admits only a finite number of solutions $h = k + 1$ for which $h + l_j = 0$ there must be different complex fractions satisfying the same equation (2) with $-k_j > 0$. These equations have only one positive root. Therefore the same complex fraction will be repeated. Hence the original fraction is 1.

The period of a continued fraction will be denoted by a bar, i. e. j [4/2]

$$(a_1, \dots, a_m, \overline{a_{m+1}, \dots, a_{m+j}}) \quad (3.7)$$

denotes a continued fraction with a period j , $a_{m+j} = a_{m+1}$. If $j = 0$ the continued fraction is said to be periodic.

Let a_1, a_2, \dots be the complete fraction $\frac{1}{a_1 + \frac{1}{a_2 + \frac{1}{a_3 + \dots}}}$. As $a = \frac{1}{a_1 + \frac{1}{a_2 + \frac{1}{a_3 + \dots}}}$ every property which holds for simple fractions of sufficient high index holds for every complete fraction of index n . From (3.4) we know that the real part of a n -th order fraction of a n -th order high index is negative. Therefore the property holds for every complete fraction of index n or $n-1$ for every purely periodic fraction. In a purely periodic continued fraction a_1, a_2, \dots hence the continued fraction represents a number > 1 .

A root λ of a quadratic equation $\lambda^2 + p\lambda + q = 0$ is called λ if $\lambda > 1$ and the conjugate root satisfies $|\lambda| > 1$ and $|\lambda| < -1$. We will prove now that every purely periodic continued fraction represents a reduced quadratic number.



Let

$$\alpha = (\alpha_1, \dots, \alpha_n) = (\alpha_1, \dots, \alpha_n | \alpha) \quad (3, 7)$$

$$\xi = (\xi_1, \dots, \xi_1) = (\xi_1, \dots, \xi_1 | \xi) \quad (3, 7')$$

be two purely continued fractions, the elements α_i being the same in both continued fractions, but ordered in an inverse manner.

Let P_n, Q_n be the convergents of α . Then (see 1, 17')

$$\frac{P_n}{Q_n} = \frac{P_{n-1} + \alpha_n Q_{n-1}}{Q_{n-1} + \alpha_n Q_{n-2}} = (\alpha_n, \dots, \alpha_1)$$

hence $\frac{P_n}{Q_n} = \frac{Q_n}{P_n}$ if

Let $\beta = \frac{1}{\alpha}$, then $0 > \beta > -1$, and

$$P_{n-1} + \alpha^n P_{n-2} = P_{n-1} + Q_{n-1} - 1 = 0 \quad \text{hence } \alpha \text{ is a root of}$$

$$f(x) = Q_n x^2 + (Q_{n+1} - P_n)x - P_{n+1}$$

As $\alpha = \frac{P_n}{Q_n} > 0$, α is a root of $f(x)$, and as $\alpha > 1$,

the other root $\beta = -\frac{1}{\alpha}$ is different from α . Hence we consider α as a three-term

continued fraction, and β as a purely continued fraction (1, 7).
 If α is rational, β is rational, and $\alpha + \beta = 1$. If α is irrational conjugate to α_1 and $\xi = -\beta^{-1}$, then ξ is represented by (3, 7')

Every continued fraction is equivalent to its complete fractions except the period. Hence every continued fraction is equivalent to a purely continued fraction; hence

Corollary. Every quadratic number is equivalent to a reduced quadratic number.

1. Every rational number is equivalent to a purely continued fraction. Every quadratic number is equivalent to a purely continued fraction.

$$\alpha = \frac{a + \sqrt{D}}{b} = s + \frac{1}{t}, \text{ where } s < \alpha < s + 1,$$

and $a, b, s, D > 0$ are integral numbers.

$$\frac{a + \sqrt{D}}{b} = s + \frac{1}{t} \Rightarrow \frac{a + \sqrt{D}}{b} - s = \frac{1}{t} \Rightarrow \frac{a + \sqrt{D} - st}{b} = \frac{1}{t} \Rightarrow b' = (D - a^2) : b$$

From these formulas we can find out by a simple number on which the numbers a_1, a_2, \dots defining uniquely the complete fraction α and the numbers b_1, b_2, \dots defining the primary fraction β . As this continued fraction is periodic one period must be repeated after a finite number of steps. Then the first period is finished and it would be how to be stopped.

Examples.

1. $\alpha = \frac{1 + \sqrt{5}}{2}$ (harmonic section) $D=5$

$$\begin{aligned} a_1 &= 1, & b_1 &= 1, & a_2 &= 1, & b_2 &= 1, \\ 1 &= 2, & a_3 &= 1 + \sqrt{5}, & a_4 &= 1, & b_4 &= 1, \\ 1 &= \frac{1+\sqrt{5}}{2}, & a_5 &= 1 + \sqrt{5}, & a_6 &= 2, & b_6 &= 1, \\ 1 &= 2. \end{aligned}$$

The last complete fraction is therefore equal to the preceding hence $\alpha = 0.1$. This is the simplest continued fraction the worst for practical calculation. The numbers P_n, Q_n are increasing slower than in any other case,

2. $\alpha = \sqrt{20}$.

$$\begin{array}{r} a_1 = 4, \quad b_1 = 1, \quad a_2 = 5, \\ 0 \quad \quad \quad 1 \quad \quad \quad 5 \\ \quad \quad \quad 1 \quad \quad \quad 10 \\ 5 \quad \quad \quad 1 \end{array}$$

hence $\alpha \approx 4.47$
calculation

This example is very convenient for quick and exact

$P_0 =$	1	$Q_0 =$	0
$P_1 =$	5	$Q_1 =$	1
$P_2 =$	9	$Q_2 =$	10
$P_3 =$	41	$Q_3 =$	101
$P_4 =$	121	$Q_4 =$	1020
$P_5 =$	525	$Q_5 =$	10301
$P_6 =$	1401	$Q_6 =$	104030

Hence $a = \frac{5001}{10000} - \epsilon$, where $0 < \epsilon < 10^{-11}$. Therefore
 $a = 5000 P_1 - 1$ (i.e. the last two figures being uncertain).

As the error $\epsilon = \frac{1}{Q_1} - \frac{P_1}{Q_1} < \frac{1}{Q_1 Q_1 - 1} < \frac{1}{Q_1^2 - 1}$, it is useful to
 stop the calculation just before a large a_{n+1} .

$$\begin{array}{rcl} \sqrt{2} = \frac{a}{b} & D = 2 & \begin{array}{c} a \\ b \end{array} \\ \begin{array}{c} 1 \\ 1 \\ 1 \end{array} & & \begin{array}{c} 1 \\ 1 \\ 1 \end{array} \end{array}$$

Hence $\sqrt{2} = \frac{1}{1}$. As P_1, Q_1 are increasing very slowly we will use
 another method.

$$\sqrt{2} = \sqrt{200} \cdot \frac{1}{10} \quad \text{If } \sqrt{200} = \frac{P_2}{Q_2} + \epsilon \quad \sqrt{2} = \frac{P_2}{10Q_2} \pm \epsilon$$

So we represent $\sqrt{200}$ by a continued fraction $D = 200 = 14^2 + 4$

$$\begin{array}{l} \frac{a}{b} = \frac{14}{1} + \frac{4}{14} < \sqrt{200} < \frac{15}{1} \\ 14 + \frac{4}{14} < \sqrt{200} < 15 \\ 14 + \frac{1}{4} < 14 + \sqrt{200} < 20 \end{array}$$

Hence $\sqrt{200} = (14, 7, 28)$,

$$\begin{array}{ll} P_0 = 1 & Q_0 = 1 \\ P_1 = 14 & Q_1 = 1 \\ P_2 = 90 & Q_2 = 7 \\ P_3 = 2794 & Q_3 = 197 \\ P_4 = 19601 & Q_4 = 1393 \end{array}$$

$$\sqrt{200} = \frac{19601}{1393} - \epsilon, \quad 0 < \epsilon < \frac{1}{Q_4 Q_3} < \frac{1}{28 Q_4} < 3 \cdot 10^{-6}$$

$$\sqrt{2} = \frac{19601}{1393} - \epsilon', \quad 0 < \epsilon' < 3 \cdot 10^{-6} = 1.414213561$$

true to eight figures after the decimal point

Exercise. Prove that $\alpha(20) = \sqrt{41} + 1$, and calculate $\sqrt{2} + 1$, $\sqrt{13}$, $\sqrt{5} + \frac{1}{3}$, $\sqrt{17}$. Calculate $\sqrt{3}$ directly and also by help of $\sqrt{20}$.

In order to prove the converse of the last theorem, it is useful to consider the following lemma.

Lemma. If $\alpha > 1$ and $\beta < 0$ are conjugate quadratic numbers $\alpha = (a, a_1, a_2, \dots)$, then all the complete fractions $\alpha_1 = (a_1, a_2, \dots)$, $\alpha_2 = (a_2, \dots)$, are reduced numbers.

Proof. $\alpha = a + \frac{1}{\alpha_1}$, $\beta = a + \frac{1}{\beta_1}$, α_1 and β_1 are conjugate numbers

$\alpha_1 > 0$, $\frac{1}{\alpha_1} = \alpha - a > a - 1$ hence α_1 is reduced, and by repetition of this procedure it follows that α_2, \dots are reduced.

Theorem. Every reduced quadratic number is represented by a purely periodic continued fraction.

Proof. Every quadratic number is represented by a periodic continued fraction $\alpha = (a, a_1, \dots, a_n)$. Let this number be reduced and let the periodicity of the continued fraction begin with a_1 only, i.e. let $a \neq a_n$, then it follows from the last lemma that $\alpha = (a, a_1, \dots, a_n)$ is a reduced number (1). We will prove that this is impossible. Using the same notations as in the lemma we state

$$\alpha_1 = \alpha_{n+1}$$

$$\text{hence } \beta_1 = \beta_{n+1}$$

$$\alpha = a + \frac{1}{\alpha_1}, \quad \alpha_n = a_n + \frac{1}{\alpha_{n+1}}, \quad \text{hence } \beta = a + \frac{1}{\beta_1}, \quad \beta_n = a_n + \frac{1}{\beta_{n+1}}$$

$$\frac{1}{\beta_1} = \alpha - \beta, \quad \frac{-1}{\alpha_{n+1}} = \alpha_n - \beta_n,$$

but as α and α_{n+1} are reduced $0 < \alpha - \beta < 1$ and $0 < -\beta_n < 1$

hold, hence $a - 1 < \frac{1}{\beta_1} < a$, and $a_n - 1 < \frac{-1}{\beta_{n+1}} < a_n$. From $\beta_1 = \beta_{n+1}$

it follows therefore that $a = a_n$.

Theorem. Let $\alpha = \sqrt{\frac{c}{d}}$, $c > 1$ be rational then

[9/5]

$$\alpha = (a, a_1, \dots, a_n)$$

(A, B)

$$\text{and } a_i = 2a$$

$$(i = 1, \dots, n-1) \quad (8, 8')$$

$$a_i = a_{i+1}$$

hold. If conversely $a_i = a_{i+1}$ and $a_i \neq a_{i+1}$ hold, then a_i is an irrational square root > 1 .

Proof. As $a_i > 1$ and the number $a_{i+1} < 0$ is conjugate to a_i , it follows from the lemma that the numbers a_1, a_2, \dots, a_{i-1} are reduced and therefore $a_{i-1} = 1$. Hence a_i satisfies $a_i^2 = a_{i-1}^2 = 1$,

$$\text{and } a_i = a_{i+1} = a_i + \frac{1}{a_i}, \text{ and } a_i \text{ is conjugate to } a_{i+1}$$

$$a_i = (a_1, \dots, a_n) \quad (8, 9)$$

Therefore it follows from the remark (2) that

$$a_i = 1 : \beta_i = (a_n, \dots, a_1) \quad (8, 9')$$

$$0 = a_i + \beta_i = (a_i + a_1 + 1) \dots \quad \text{Hence } a_i = 1 : \beta_i = a_i + a_1$$

$$\text{Therefore } (a_i, \beta_i) = (2a_i, \frac{1}{a_i}) \quad (8, 10)$$

It is easily seen that (8, 10) is equivalent to (8, 10) if a_i is defined by (8, 8), (8, 8'), then (8, 10) holds.

Let a_1 and a_2 be defined by (8, 8) and (8, 8'), and let $\beta = a_1 + 1 : a_2$, then it follows from (8, 9) and (8, 9') that a_1 and a_2 and therefore β are conjugate numbers. It follows that $a_1 + \beta = 0$. Hence a_1 and β are the roots of a rational polynomial $x^2 + 0x - 1 = 0$. From $0 \neq a_1 = 2a_i$ it follows that $a_i \neq 1$, and therefore $a_i > 1$ and as (8) is an infinite continued fraction, a_i must be irrational.

Corollary. Let $a = p/r$ and $P_1 = Q_1$ be the convergents of a , then

$$(P_1^2 - rQ_1^2) = (-1)^{k+1}r \quad (8, 11)$$

holds for every $k = 1, 2, \dots$

Proof. Let a_i be the complete fractions of a , then $a_i = a_{i+1} +$

$$a_{i+1} = a_{i+1} + \frac{1}{a_{i+1}} = 2a_i + \frac{1}{a_i} = a_i +$$

$$\frac{1}{a_i} = a_i +$$

$$\text{But as } a = \frac{P_{2n-1} + P_{2n-2}t}{Q_{2n-1} + Q_{2n-2}t}, \quad (3, 13)$$

$$a = \frac{P_{2n}(s + a) + P_{2n-1}}{Q_{2n}(s + a) + Q_{2n-1}}, \quad \text{hence, hence}$$

$$Q_{2n}as - P_{2n}s = P_{2n-1} + a(Q_{2n}s + Q_{2n-1}) - P_{2n-1} = 0$$

As $s = r + t$ is rational, and a is irrational

$$\begin{aligned} P_{2n-1} + P_{2n-2}t &= Q_{2n-1} + Q_{2n-2}t \\ &= P_{2n} + Q_{2n}s + Q_{2n-1} = 0 \end{aligned}$$

hold. If we multiply these equations by $Q_{2n}t$ respectively $-P_{2n}t$ and add, we get $(P_{2n}^2 - tQ_{2n}^2 + t(P_{2n-1}Q_{2n} - Q_{2n-1}P_{2n})) = 0$ and from this formula we get (3, 11) directly.

§ 4. APPLICATIONS OF THEORY OF NUMBERS

It is proposed to solve

[4, 1]

$$x^2 - dy^2 = 1 \quad (4, 1)$$

by integral x and y

Obviously (4, 1) cannot be solved if there is a common factor r of a and b different from ± 1 . We therefore suppose a and b to be relatively prime. a/b can be represented by an even continued fraction $a/b = [2, \dots]$

$$a/b = (p_1, \dots, p_{2n})$$

$$a/b = P_{2n}/Q_{2n}, \text{ and as } a \text{ and } b \text{ are positive and relatively prime}$$

$$a = P_{2n}, b = Q_{2n} \text{ and therefore}$$

$$aQ_{2n-1} - bP_{2n-1} = P_{2n}Q_{2n-1} - Q_{2n}P_{2n-1} = (-1)^{n-1}$$

holds. Hence we get the integral solutions by

$$x = Q_{2n-1} + k b,$$

$$y = P_{2n-1} + k a$$

where $k = 0, \pm 1, \pm 2, \dots$

To solve $x^2 + dy^2 = 1$ Pell's equation by integral x and y we will use (4, 2) (3, 11).

$$\sqrt{d} = a = a_0, a_1, \dots$$

$$\text{then } P_{2n}^2 - dQ_{2n}^2 = (-1)^{n-1}$$

Therefore if n is even, $(x, y) = (P_{2n}, Q_{2n})$

and if n is odd, $(x, y) = (P_{2n+1}, Q_{2n+1})$

are solutions for every positive integer n .

$$K[y]: x^2 - 20y^2 = 1$$

$$\sqrt{20} = (5, 10) \quad n = 1$$

By this method we got the solutions

$$\begin{aligned} (x, y) &= (P_2, Q_2) = (51, 10) \\ &= (P_4, Q_4) = (521, 1021) \\ &= (P_6, Q_6) = (53851, 108030) \end{aligned}$$

4.5. CONTINUED FRACTIONS AND FERMAT'S CHL $\phi(x)$

[1] In Unit 4 the system of the positive numbers has been taken for the system $A = \mathbb{N}$ being the ring of the integers. We will now consider another system A .

Let K be an arbitrary field and x an indefinite unit included in K . The elements of K will be denoted by a, b, \dots, t with and without a bar. The elements of the ring $K[x]$ will be denoted by

$$f(x), g(x), \quad (5, 1)$$

(the ring is denoted of a field) $K[x]$ will be set as the ring S .

In order to get a new system A we create new elements denoted by Greek letters

$$\phi(x), \psi(x), \chi(x), \omega(x) \quad (5, 2)$$

in the following manner.

$$\begin{aligned} \phi(x) &= a_n x^n + a_{n-1} x^{n-1} + \dots + a_0 + a_{-1} x^{-1} + \dots + a_{-n} x^{-n} + \dots + \\ &= (a_n x^{n-n} + \dots + 0) x^{n-1} + \dots + a_n x^0 + \dots + a_{-n} x^{-n} + \dots = \sum_{i=-\infty}^{\infty} a_i x^i \end{aligned} \quad (5, 3)$$

This is a pure & formal definition. It means that to every sequence of coefficients from K with fixed decreasing integral indices

$$a_n, a_{n-1}, \dots, \quad (5, 4)$$

then φ is one of our new elements and this element will not be changed if we take before φ a finite set of numbers. Nothing has been supposed about convergence. We have to define the addition and the multiplication of the elements (5.2) and we will let to them in such a way that the elements (5.2) for which $k = 0, \infty$ (or $k < 0$) form a subring isomorphic to $K[x]$.

So we define :

$$\text{Let } n \geq m, \varphi(x) = \sum_{i=-n}^{-m} a_i x^i$$

$$\psi(x) = \sum_{i=-r}^{-k} b_i x^i = \sum_{i=-r}^{-k} c_i x^i \quad \text{where } c_i = b_i, \quad i = -r, \dots, -n-m$$

$$\text{then } \varphi(x) + \psi(x) = \chi(x) = \sum_{i=-n}^{-m} c_i x^i$$

$$\varphi(x) \cdot \psi(x) = u(x) = \sum_{i=-n}^{-r} d_i x^i$$

$$\text{where} \quad c_i = a_i + b_i, \quad d_i = \sum_{j+k=i} a_j b_k, \quad (5.4)$$

$$n \geq i \geq k-m.$$

The definitions (5.4) are obviously independent of null coefficients put before, the commutative, associative and distributive laws hold, and the subtraction is uniquely defined by

$$b_i = a_i - c_i.$$

Hence for the null element every coefficient will be 0. If for the elements of $\varphi(x)$ the coefficients $b_i = 0$, for $i \leq k$ hold then $n - n - \varphi(x) - \psi(x) = \sum_{i=-n}^{-m} c_i x^i$

$$d_i = b_i \cdot a_i \quad \text{holds.}$$

The elements (5.2) form a ring K and those elements for which $k = 0, \infty$ form a subring for which the addition and multiplication has been defined in the same way as for polynomials. Hence there is an isomorphism f by which this subring becomes isomorphic to $K[x]$. Let $\varphi(x) \neq 0$, then $\varphi(x)$ has at least one coefficient $\neq 0$, let n be the highest index of the non vanishing coefficients, then

$$\varphi(x) = \sum_{i=-n}^{-m} a_i x^i, \quad a_n \neq 0.$$

n is an integer the degree of x^n . From (2) it follows directly

The degree of a product is equal to the sum of the degrees of the factors.

The degree of a sum of elements of different degrees is equal to the maximum degree of the summands.

From the remark at the end of §1 it follows that a product of two elements $\neq 0$ cannot be equal to 0. Hence the ring of the elements of R is an integral domain. We will now consider the elements of $K[x]$ with the corresponding elements of R .

So the elements $\sum_{j=0}^{\infty} c_j x^j$, $c_j \in R$, $c_0 \neq 0$ are identified with b_0

and $b_0 \phi(x) = \sum_{j=0}^{\infty} b_j a_j x^j$ holds

Let $\phi(x) = \sum_{j=0}^{\infty} c_j x^j$, $c_j \in R$, $c_0 \neq 0$. Then $x^k \phi(x)$, $k \geq 1$. Every field F containing R contains no zero divisors of R . The elements of R , which are $\neq 0$ are elements of K . K is a ring which is isomorphic to a subring of F . The elements of this ring will therefore be identified with the corresponding elements of F . Since $\phi(x)$ is nonzero identified with ϕ and the same sum $\sum_{j=0}^{\infty} c_j x^j$ becomes identified with the symbol

sum $\phi(x) = \sum_{j=-m}^{\infty} a_j x^j$, where $a_j = 0$, for $j < -m$.

Using these notations we can extend the operations of division of the polynomials to the elements of R .

Let $\phi(x)$ and $\psi(x)$ of degree n and m respectively, a_n and b_m their highest coefficients.

$$\frac{a_n}{b_m} \phi(x) = \phi_1(x) = \phi(x) - \frac{a_n}{b_m} \psi(x) \text{ is of degree } n_1 < n$$

By repetition of this procedure we get

$$\frac{a_n}{b_m} \phi(x) = \phi_1(x) + \frac{a_n}{b_m} \psi(x) = \phi_2(x) + \frac{a_n}{b_m} \psi(x) + \frac{a_n}{b_m} \psi(x) + \dots + \frac{a_n}{b_m} \psi(x) + \frac{a_n}{b_m} \psi(x)$$

and by further repetition we get an enumerable set of elements c_i of K

$$i \leq n - n_1 \text{ such that } \frac{a_n}{b_m} \phi(x) = \sum_{i=0}^{n-n_1} c_i x^i \text{ and}$$

$$\phi_1(x) = \phi(x) - \frac{a_n}{b_m} \psi(x) = \phi_2(x) \text{ is of degree } n_2 < n_1$$

$$\text{Let } \chi(x) = \sum_{j=0}^{\infty} c_j x^j = \frac{a_n}{b_m} \phi(x) + \frac{a_n}{b_m} \psi(x)$$

then $\phi_1(x)$ is of degree $\leq n_1 - n_2 + 1$ and $\phi_1(x) \neq 0$ since $\phi(x)$ is of degree $\leq n_1 - 1$.

$\phi(x) = \psi(x)\lambda(x)$ where $\lambda(x) = \lambda_1(x) + \lambda_2(x) + \dots + \lambda_n(x)$ and $\deg \lambda_i \leq n_i$ for every i , hence $\deg \lambda \leq n$. Hence $\phi = \psi \lambda$, $\deg \psi \leq n - 1$. As $\psi(x)$ was supposed to be an arbitrary element of \mathcal{Q} , we have

Theorem. The set of all elements of \mathcal{Q} is a field containing the quotientfield \mathcal{Q} of $K[x]$.

$$f(x) \phi(x) = \sum_{i=1}^n a_i x^{n_i} = \sum_{i=1}^n a_i x^{n_i} + \sum_{i=1}^n x^{n_i} \phi_i(x) = f(x) + \phi_1(x) \quad [5.2]$$

This representation of ϕ as a sum of polynomials $a_i x^{n_i}$ and an element which has degree of $\phi_1(x) \leq n_1 - n_2 + 1$.

As $1/\phi_1(x)$ is of degree ≥ 0 (since $n_1 \geq n_2$), it follows that there is one and only one representation $\phi = \psi(x) + \phi_1(x)/\phi_2(x)$ or $\phi(x) = f(x) + \lambda(x)/\mu(x)$, where the degree of $\lambda(x) < \deg \mu(x)$.

So we can apply (3.1) to (5.1) and to the case when the elements

$$x^{n_1}, x^{n_2}, \dots, x^{n_r}$$

of \mathcal{A} are the elements $x^{n_1}, x^{n_2}, \dots, x^{n_r}$ and $\lambda_1, \lambda_2, \dots, \lambda_r$ are the polynomials in x . [See (1, 1), (1, 2)]

If we make this assumption about \mathcal{A} and \mathcal{B} , then $\lambda_1, \lambda_2, \dots, \lambda_r$ are uniquely defined by (4.1) and serve as unique representatives of the elements of \mathcal{A} as contained in \mathcal{B} . The λ_i are polynomials in x of degree ≥ 0 . If we put $\lambda_i(x) = \lambda_i$ in (4.1), then λ_i is a polynomial of degree ≥ 0 or the next element of \mathcal{A} is a polynomial. Representation of every element of \mathcal{A} as contained in \mathcal{B} is then

Theorem. The elements of \mathcal{A} can be represented in one and only one manner by continued fractions (c_1, c_2, \dots, c_n) where c_i is a polynomial in x , whose degree is ≥ 0 , for $i \geq 1$.

The degree of a polynomial λ has just the same properties as the function $\deg \lambda$ [1, 2]. From that and (4.1) it follows that we get therefore the

Corollary. The finite continued fractions represent the elements of \mathcal{Q} and every element of \mathcal{Q} is represented by a finite continued fraction.

We will now use the representation of real numbers by continued fractions in order to represent the elements of \mathfrak{A} .

[5/8] As (1, 15) holds,

$$a_1 - \frac{P_1}{Q_1} = \frac{1 + (-1)^{k-1} a_{k+2}}{a_2 Q_1}$$

we will prove that the right side of this equation is an element whose degree decreases indefinitely as k increases.

The elements a_i have been defined by $a_i = c_i + d_{i+1}$, a common factor being arbitrary (see (1, 12)) hence $a_2 = a_1 + 1 = c_1 + d_2$; therefore

$$\text{degree} \left(a_1 - \frac{P_1}{Q_1} \right) = -\text{degree } Q_1 - d_2 - \dots = -1. \quad (5, 5)$$

$d_i > 0$ for $k > 1$ and $a_i = c_i + 1 + c_{i+1}$ (see 1, 4). As the degree of a sum of s integers of different degrees is equal to the highest of the degrees of the summands $\text{degree}(c_i) = \text{degree}(c_i) = d_i$ for $i > 1$

$$Q_1 = 1, Q_2 = a_2, \dots, Q_i = a_i Q_{i-1} + Q_{i-2}.$$

Hence $\text{degree } Q_i > 1$ for $i > 1$, hence the degrees increase with the index and therefore

$$\text{degree } Q_i = \text{degree}(a_i Q_{i-1}) = i + \text{degree } Q_{i-1} = \sum_{j=1}^i d_j. \quad (5, 6)$$

Hence from (5, 5) and (5, 6) it follows that

$$\text{degree} \left(a_1 - \frac{P_1}{Q_1} \right) = -1 = -2 - \sum_{j=2}^k d_j - d_{k+1} =$$

$$\text{degree} \left(\frac{-1}{Q_1 Q_{k+1}} \right) \leq 1 - 2k. \quad (5, 7)$$

The efficiency of the approximation of an real number α by a continued fraction becomes clear by the theorem that if $\frac{p}{q}$ approximates α

better than $\frac{P_1}{Q_1}$, then $b > Q_1$ holds. The corresponding theorem holds

in the case we consider here.

Theorem. If $f(x)$ and $g(x)$ are polynomials of $K[x]$

$$d = \frac{f(x)}{g(x)} = \frac{P_{d-1}}{Q_{d-1}} \neq 0 \quad \text{and}$$

$$r = \text{degree} \left(r_1 - \frac{f(x)}{g(x)} \right) < d = \text{degree} \left(r_1 - \frac{P_{d-1}}{Q_{d-1}} \right), \text{ then}$$

$$\text{degree } g(x) > \text{degree } Q_{d-1}, \text{ holds} \quad (5.6)$$

Proof. $f(x) = \left(r_1 - \frac{P_{d-1}}{Q_{d-1}} \right) \cdot \left(\frac{f(x)}{g(x)} + r_1 \right)$ The two summands on the right

side have different degrees hence $r = \log r_1$ holds. As $f(x)/Q_{d-1}(x)$ is a polynomial, $0 \leq d = \text{degree } Q_{d-1}(x) = \text{degree } f(x) + \text{degree } (Q_{d-1}^{-1} \cdot f)$

$$\text{degree } f(x) \leq r = \text{degree } Q_{d-1} - d > r = \text{degree } Q_{d-1} - d = \text{degree } Q_{d-1} - d_{d+1} > \text{degree } Q_{d-1} \text{ [see (5.7), (5.8)]}.$$

This theorem enables us to approximate functions given by a power series of $\frac{1}{x}$ by rational function in the neighbourhood of $x = 1/0$

Exercise. $\frac{1}{2} \log \frac{x+1}{x-1} = x^{-1} - \frac{1}{8} x^{-3} + \frac{1}{6} x^{-5} + \dots$

Represent this function by a continued fraction and approximate it by rational functions.

Lemma. Let $a = (a_1, \dots)$, $a' = (a'_1, \dots)$, let m be the lowest index [5/4] for which $a_m \neq a'_m$ holds and $(a_1, \dots, a_{m-1} = A, (a_1, \dots, a_{m-1} = A'$ then

$$\text{degree}(a-a') = \text{degree}(A-A') \quad \text{holds} \quad (5.9)$$

Proof. Without loss of generality we suppose that $\text{degree } a_m = r \geq \text{degree } a'_m = r'$. We shall use the ordinary notations for the convergents of a and for those of a' we shall use them with a dash,

$$P_i = 1^{\circ},$$

$$Q_i = Q'_i \quad \text{for } i < m$$

$$Q_m = a_m Q_{m-1} + Q_{m-2} \quad \text{degree } Q_m = r + q$$

$$Q'_m = a'_m Q_{m-1} + Q_{m-2} \quad \text{degree } Q'_m = r' + q$$

$$A - A' = \begin{pmatrix} P' & & 1 \\ Q' & & 0 \end{pmatrix} = \begin{pmatrix} P' & & 1 \\ Q' & & 0 \end{pmatrix} + \begin{pmatrix} 0 & & 0 \\ 0 & & 0 \\ 0 & & 0 \end{pmatrix} \\ = (-1)^{n-r} \frac{1}{Q_{n-r} Q'_{n-r}}.$$

1. If $r = r'$, $\text{degree}(p_n - p'_n) \geq 0$

$$\text{degree}(A - A') = \text{degree} \left(\frac{1}{Q_{n-r} Q'_{n-r}} \right) = -2r.$$

2. If $r > r'$, $\text{degree}(p_n - p'_n) = r$

$$\text{degree}(A - A') = \text{degree} \left(\frac{1}{Q_{n-r} Q'_{n-r}} \right) = -2r + r = -r. \quad \text{Hence}$$

$$\text{degree}(A - A') \geq \text{degree} \frac{1}{Q_{n-r}} \quad (5, 10)$$

holds in every case

From (5, 10) it follows that

$$\text{degree}(A) = \text{degree} \frac{1}{Q_{n-r}} \leq \text{degree} \frac{1}{Q_{n-r} Q'_{n-r}} = \text{degree} \frac{1}{Q'_{n-r}}.$$

$$\text{and that } \text{degree}(A - A') = \text{degree} \frac{1}{Q_{n-r} Q'_{n-r}} \leq \text{degree} \frac{1}{Q'_{n-r}}. \quad \text{Hence}$$

$$\text{degree}(A) = \text{degree} \left[\frac{A - A'}{A - A'} \right] = \text{degree}(A - A') + \text{degree}(A - A') = \text{degree}(A - A')$$

and the degree of the first term of the second summand is greater than the degree of the two other summands.

Theorem. Let x_1, x_2, \dots, x_n be a finite set of n symmetric functions of $K(x_i)$ and let $r > 1$, $\text{degree } r > 0$ so there exists a continued fraction

For the set x_1, x_2, \dots, x_n uniquely the values

$$P_1 = Q_1 \quad \text{and} \quad P_N = Q_N = (x_1, \dots, x_n)$$

Let $1 < i < N$

From the preceding lemma it follows that $\text{degree } P_N = Q_N = P_N = Q_N$

$$= \text{degree}(P_{N+1} - Q_{N+1} - P_N - Q_N) = \text{degree} \left(\frac{1}{Q_N - Q_{N+1}} \right) = -k_N.$$

where k_n increases to infinity, with n

$$P_n - Q_n \sum_{k=0}^{\infty} a_k x^k = \sum_{k=0}^{k_n-1} b_k x^k + \sum_{k=k_n}^{\infty} c_k x^k \quad (5, 11)$$

The coefficients b_k are independent of n . As k_n increases with the index n we get an infinite set b_0, b_1, \dots defining

$$\phi(x) = \sum_{k=0}^{\infty} b_k x^k$$

Finally we have to prove that $\phi(x) = (s_1, s_2, \dots)$. Let $\phi(x) = (s_1, s'_2, \dots)$, and m be the smallest index for which $s_m \neq s'_m$. Then it follows from the lemma that for every $n \geq m$,

$$\text{degree } (\phi(x) - (s_1, \dots, s_n)) = \text{degree } (\phi(x) - (s_1, \dots, s_n)) \text{ holds.}$$

But $\phi(x) - (s_1, \dots, s_n) = b_{k_n} x^{k_n} + \dots + b_{k_n+1} x^{k_n+1} + \dots$ is of degree k_n and decreases infinitely with n .

6. CONTINUED FRACTIONS WITH RATIONAL ELEMENTS

Let \mathbb{R} be the field of the rational numbers. Then every finite continued [6, 1] fraction

$$\begin{aligned} (s_1, \dots, s_n) &= \frac{P_n}{Q_n} = \frac{P_1}{Q_1} + \left(\frac{P_2}{Q_2} - \frac{P_1}{Q_1} \right) + \dots + \left(\frac{P_n}{Q_n} - \frac{P_{n-1}}{Q_{n-1}} \right) \\ &= s_1 + \frac{1}{Q_1 Q_2} + \dots + \frac{1}{Q_{n-1} Q_n} \end{aligned} \quad (6, 1)$$

represents a rational number, but an infinite continued fraction defines a number if and only if

$$\sum_{k=1}^{\infty} \frac{(-1)^k}{Q_{k-1} Q_k} \quad (6, 2)$$

converges. If the sum (6, 2) is convergent

$$(s_1, s_2, \dots) \quad (6, 3)$$

defines a real number equal to (6.3). A necessary condition for the convergence of (6.3) is therefore

$$|Q_{n-1}Q_n| \longrightarrow \infty \quad (6.4)$$

If the numbers q_n are either > 1 or < 1 each q_n of alternating sign, the sum (6.2) is an alternating sum. Hence the continued fraction converges if $|Q_n|Q_{n-1}|$ increases steadily and (6.4) is satisfied.

[6/2] **Theorem.** If $\sum |q_n|$ is convergent, (6.3) is divergent.*

Proof. We prove by mathematical induction that

$$Q_n < \prod_{j=1}^n (1 + |q_j|) \quad (6.5)$$

As $Q_1 = 1$, $Q_2 = q_2$, the formula holds for $n < 3$. If (6.5) is true for $n < m$,

$$\begin{aligned} Q_m &= q_m Q_{m-1} + Q_{m-2} \\ |Q_m| &\leq \prod_{j=1}^m (1 + |q_j|) = \left(\prod_{j=1}^{m-1} (1 + |q_j|) + 1 \right) \\ &< \prod_{j=1}^m (1 + |q_j|). \end{aligned}$$

If $\sum q_n$ converges, the infinite product $\prod (1 + |q_n|)$ converges to a positive number Q and $|Q_n| < Q$ holds for every index n . Hence (6.4) does not hold and the continued fraction is divergent.

[6/3] Let $q > 1$ for $n > 1$. If in $Q_1 = 1$, $Q_2 = q_2 > 0$, $Q_n = q_n Q_{n-1} + Q_{n-2}$ it follows by mathematical induction that each number $Q_n > 0$.

$$Q_n Q_{n-1} = q_n Q_{n-1}^2 + Q_{n-1} Q_{n-2} > Q_{n-1} Q_{n-2} > 0$$

$Q_n Q_{n-1}$ is therefore an alternating series whose elements have steadily decreasing absolute values. This series converges therefore if and only if (6.4) is satisfied. These considerations lead to the following

Theorem. Let $q_n > 1$ for $n > 1$. Then the continued fraction (6.3) is convergent if and only if $\sum q_n$ is divergent.

Proof. If $\sum q_n = \sum |q_n|$ is convergent, the continued fraction is divergent, as it has been proved by the preceding theorem.

* We use the term "divergent" for every non-convergent series.

Let $\sum s_i$ be divergent then $s_i \rightarrow \infty$. As $Q_0 > 0, Q_1 = 1$, and $Q_{2n+1} = s_{2n+1}, Q_{2n+2} = Q_{2n+1}$ we see by mathematical induction that $Q_{2n+1} \geq 1$. Hence $Q_{2n+2} = s_{2n+2} Q_{2n+1} + Q_{2n+2} \geq s_{2n+2} + Q_{2n+2}$ and as $Q_0 = s_0$, it follows by mathematical induction that

$$Q_{2n} \leq \sum_{i=0}^n s_i$$

Hence

$$Q_{2n+1}, Q_{2n+2} \rightarrow \infty, \text{ and } Q_{2n} - Q_{2n-1} \rightarrow \infty$$

therefore (6.4) is satisfied and as we stated above this condition is sufficient for the convergence of (6.3) in the case considered here.

Let $s_1 = 1, s_i \geq 1$ for $i > 1$ then $s_i \rightarrow \infty$ and it follows from the (6.4) preceding theorem that the continued fraction (6.3) converges to a value in the interval

$$\left(\frac{P_1}{Q_1} + \frac{P_0}{Q_0} \right) = (0, 1)$$

We will show that this value is irrational.

Let $a_1 = (s_1, s_2, \dots)$ be rational say $a_1 = \frac{a_1}{a_2}$ where a_1, a_2 are integral

then $a_1 > a_2$ and $\frac{1}{a_1} = \frac{1}{s_1 + s_2}$ hence $s_2 = \frac{1}{a_1} - s_1 = \frac{a_2}{a_1}$, where $a_3 = a_1 - a_2 a_2$ is

integral as $a_2 = (s_2, s_3, \dots)$, there $0 < s_2 < 1$ hence $a_2 > a_1 > 0$. In the same manner we get

$$a_3 = \frac{1}{a_2 + s_3}, \dots, a_4 = \frac{1}{a_3 + s_4} = \frac{a_4}{a_5} \text{ and } a_1 > a_2 > 0$$

By repetition of this procedure we get an infinite set of decreasing integral positive numbers

$$a_1 > a_2 > a_3 > a_4 > \dots$$

and that is impossible.

Hence a_1 is irrational.

Example $s_1 = 0, s_2 = 2, s_3 = 1$ and for $n \geq 1$

$$s_n = \frac{2 \cdot 4 \cdot 2n + 2}{2n + 1} > 1, s_{n+1} = \frac{3 \cdot 4 \cdot 2n + 1}{2 \cdot 4 \cdot 2n + 2} > 1$$

then it is

$$Q_1 = 1, Q_2 = 2, Q_3 = 3$$

and from the identity

$$\begin{aligned} 2 \cdot 4 \cdot 2m + 1 &= 1 \cdot 3 \cdot (2m + 1) + s_{2m} + 2 \cdot 4 \cdot 2m + 2 \\ 1 \cdot 3 \cdot (2m + 1) &= 2 \cdot 4 \cdot 2m + s_{2m+1} + 1 \cdot 3 \cdot (2m + 1), \end{aligned}$$

it follows by mathematical induction that

$$Q_{2m} = 2 \cdot 4 \cdot 2m, Q_{2m+1} = 1 \cdot 3 \cdot (2m + 1)$$

hence

$$Q_n + Q_{n+1} = n!$$

the continued fraction is irrational. Its value is

$$\frac{1}{Q_1 + Q_2} = \frac{1}{Q_2 + Q_3} + \frac{1}{Q_3 + Q_4} + \dots = \sum_{n=1}^{\infty} \frac{1}{n!} = e^{-1}.$$

Hence e is irrational.



PART IV.
APPROXIMATE SOLUTION



§ 1. HORNER'S SCHEME

Let K be an arbitrary field, \mathbb{R} , the field of the real numbers or $\{1/1\}$ of the complex numbers, q be an element of K and $f(x)$ a polynomial of $K[x]$,

$$f(x) = \sum_{i=0}^n a_i x^i = (x - q) \sum_{i=0}^{n-1} a_i' x^i + a_n' x - q f_1(x) + a_0,$$

then

$$a_i = a_{i+1}' - f_1(x) a_{i+2}', \quad \text{Hence}$$

$$a_0 = a_1'$$

$$a_1 = a_2' + q a_1' \quad (1.1)$$

$$a_n = a_{n+1}' f_1(x)$$

holds. We can arrange the calculation of the coefficients a_i , a_{i+1}' as

$$\begin{array}{ccccccc} a_n & a_{n-1} & a_{n-2} & \dots & a_1 & a_0 \\ & q & & & q^2 & & \\ & a_{n-1}' & a_{n-2}' & & a_1' & a_0' & \\ \hline a_n & a_{n-1} & a_{n-2} & \dots & a_1 & a_0 \end{array} \quad (1.1')$$

$$a_n = a_{n-1}' - f_1(x) a_{n-2}', \quad a_{n-1} = a_{n-2}' - f_1(x) a_{n-3}',$$

$$f_1(x) = x - q, \quad (1.2)$$

We can find $f_1(x)$ by the same method $f_1(x) = x - q$ and so on

$$f_1(x) = (x - q) f_{11}(x) + a_n', f_{11}(x) = \sum_{i=0}^{n-2} a_{i+2}' x^{i+2}$$

After $n-1$ steps we get $f(x)$ represented as a polynomial in $x - q$

$$f(x) = a_0 + a_1'(x - q) + \dots + a_{n-1}'(x - q)^{n-1} + a_n'(x - q)^n$$

This representation is known in Analysis as the Taylor series $f(x)$ at the point $x = q$. The successive calculation of the coefficients can easily be done on using the last one of (1.1') up to a_1' . The complete scheme for this calculation is called Horner's scheme. It is the most convenient method

for calculating r , p , if q and the coefficients of $f(x)$ are given and furnishes the representation of r by $x = q + 1$. The calculation will be explained by the following

Example. $f(x) = x^5 - 13x^4 + 68x^3 - 119x^2 + 67x - 65$

$$\begin{array}{r}
 q=1. \quad \begin{array}{rrrrr}
 1 & -13 & 68 & -119 & 67 \\
 & 1 & -14 & 54 & -65 \\
 \hline
 1 & -14 & 54 & & \\
 & 1 & -13 & 41 & -61 \\
 \hline
 1 & -13 & 41 & & \\
 & 1 & -29 & 20 & -21 \\
 \hline
 1 & -12 & 20 & & \\
 & 1 & -24 & 2 & \\
 \hline
 1 & -11 & 2 & &
 \end{array}
 \end{array}$$

$$\begin{aligned}
 f(x) &= (x-1)(x^4-12x^3+20x^2-24x+2) \\
 &= (x^5-13x^4+41x^3-24x^2+2x) \\
 &= (x-12)(x-1)^3+29(x-1)^3-24(x-1)+2 \\
 &= (x-1)^4-1(x-1)^3+29(x-1)^2-24(x-1)+2
 \end{aligned}$$

[1,2] If r is a rational number, the method for calculating the roots of $f(x)$ will be explained by the example of the above example.

$$\begin{aligned}
 x &= q + 1 = 1 + 0.1 = 1.1 \quad f(x) = x^5 - 13x^4 + 20x^3 - 24x^2 + 2 \\
 f(1) &= q(0) = 2 \\
 f'(1) &= q'(0) = -24
 \end{aligned}$$

We therefore suppose that there is a r such that $f(r) = 0$ near $x = 1$. In the neighborhood of $x = 1$, $f(x) \approx -24(x-1) + 2$, $x \approx 1$ for $|x-1| < 1$. Therefore we represent $f(x)$ by a polynomial in $v = 0.1$

$$\begin{array}{r}
 v=0.1 \quad \begin{array}{rrrrr}
 1 & -13 & 68 & -119 & 67 \\
 & +0.1 & -1.09 & +2.791 & -2.1209 \\
 \hline
 1 & -12.9 & 66.91 & -116.209 & 64.8791 \\
 & +0.1 & -1.28 & +2.663 & -2.1209 \\
 \hline
 1 & -12.8 & 65.63 & -114.546 & 62.7582 \\
 & +0.1 & -1.27 & +2.536 & -2.1209 \\
 \hline
 1 & -12.7 & 64.36 & -112.882 & 60.6373 \\
 & +0.1 & -1.26 & +2.409 & -2.1209 \\
 \hline
 1 & -12.6 & 63.10 & -111.218 & 58.5164
 \end{array}
 \end{array}$$

* The value $\sqrt[n]{n}$ increases approximately as $\frac{1}{n} \log n$. We apply here an elementary theorem on real number functions which will be cited in § 2.5.

As $f(1.1) = 0.1209$ there is a root between 1 and 1.1. We approximate therefore $f(x)$ by

$$18.526(x-1.1) = 0.1209, \quad \text{hence } x = 1.107 = 0.007$$

$$\begin{array}{r} q = 1.007 \quad 1 \quad 10.07 \quad 2 \quad 7.1 \quad 18.526 \quad 12.09 \\ \quad \quad \quad 0.007 \quad + \quad 0.04249 \quad + \quad 0.1885224 \quad + \quad 0.1209 \\ \hline 1 \quad 10.007 \quad 25.85421 \quad -18.70683741 \quad 0.01047878204 \\ \quad -0.007 \quad +0.074298 \quad -0.18133829 \\ \hline 1 \quad 10.014 \quad 2 \quad 5.0817 \quad 18.8812011 \\ \quad -0.007 \quad +0.074347 \\ \hline 1 \quad -19.621 \quad 25.062824 \\ \quad \quad \quad 0.007 \\ \hline 1 \quad -10.628 \end{array}$$

Hence the root is approximately equal to 1.107. The next approximation is $q = 0.0007$. If we want to obtain an even more exact value for the root the calculation would become very tedious. We therefore neglect the terms q^3 and q^4 and we get a good approximation by taking q up to q^2 . For this approximation we are left with the following equation: $18.88522017q - 0.0004788201 + 2 \cdot 0.074347q^2 - 10.628q + q^3$. As $q \approx 10^{-4}$ the two last terms will influence only the 10th and the following decimals of the equation for $-0.1 \cdot 10^{-4} < q < 0.4 \cdot 10^{-4}$ the quadratic term becomes 0.000007...

$$\text{Hence} \quad q = 0.000624 \quad x = 1.007624$$

On using this approximation we can easily get some more figures of this decimal development. The most difficult task is sometimes to get a first approximation of the roots. For this purpose it is often helpful to know the values of $f(x)$ for a suitable set of values x . We may get these values by Horner's scheme, but it is often useful to abbreviate this scheme in following manner:

Given $q_1, q_2, \dots, q_n < 1$. We calculate by Horner's scheme [1.1]

$$\begin{array}{ll} f(x) = b_1 + (x - q_1) f_1(x) & f(q_1) = b_1 \\ f_1(x) = b_2 + (x - q_2) f_2(x) & f(q_2) = b_1 + (q_2 - q_1) b_2 \\ \vdots & \vdots \\ f_{n-1}(x) = b_n + (x - q_n) f_n(x) & \\ f(x) = f_1 + x^2(x - q_1) + \dots + (x - q_1)(x - q_2) \dots + b_n(x - q_1) \dots (x - q_{n-1}) & \end{array}$$

We will follow the same procedure as in the previous example in this manner.

$$\begin{array}{rcccccc}
 q_1=1 & 1 & -15 & 08 & -119 & 67 \\
 & & 1 & -14 & 54 & -67 \\
 q_2=2 & & & & & \\
 & 1 & 11 & 2 & 1 & 2 \\
 & & 2 & -4 & 6 & \\
 q_3=3 & & & & & \\
 & 1 & -12 & 2 & & \\
 & & 3 & -2 & & \\
 & & & & & \\
 & 1 & 0 & 0 & &
 \end{array}$$

$$f(x) = (x-2)(x-1)(x+1)(x-3)(x-2)(x-3)(x-9),$$

$$f(0) = -67, f(1) = 2, f(2) = -6, f(3) = -2, f(4) = 248, f(5) = 117, f(6) = 13.$$

This expression is satisfied if $x < 1$, $x > 1$, for each of the terms becomes > 0 and that for $x > 9$, $x > 0$, for the 1st, the 4th and the sum of the two other terms becomes > 0 . So all roots are situated in the interval $(1, 9)$. We can establish no root in the interval $(1, 2)$ there is at least one root in the interval $(3, 4)$ $(-6 < -2 < 7 < 6 = 5 < -117)$. Hence there is a root in the interval $(4, 6)$. The roots may be obtained by Horner's scheme as an exercise.

[1.4] We will use here a different method of calculation. If $f(x) = \sum_{n=0}^{\infty} a_n x^n = 0$, $\frac{1}{x}$

satisfies the condition $\sum_{n=0}^{\infty} a_{n+1} \left(\frac{1}{x}\right)^n = 0$, and to every root of x in the interval

$(0, 1)$ there corresponds a value of $\frac{1}{x} > 1$. These considerations lead to

the following method of approximation due to Legendre

If ξ is a root of $f(x)$, $a < \xi < a+1 = 1 + \frac{1}{\eta_1}$, $f(x) = 0$, $x = \eta_1 \left(1 + \frac{1}{\eta_2}\right)$,

$\eta_1 > 1$ is a root of f and $b \leq \eta_1 < b+1 = \eta_1 + \frac{1}{\eta_2}$. By repetition

of this procedure we will get a representation of ξ as a continued fraction

If we stop the calculation after n steps we get the approximation $\frac{P_n}{Q_n}$

and the error becomes

$$\xi - \frac{P_n}{Q_n} < \frac{1}{Q_n Q_{n+1}} < \frac{1}{Q_n^2}$$



This method will be illustrated by the example previously used. We know that $x^4 - 15x^3 + 68x^2 - 117x + 87$ has a root in the interval $(8, 9)$. Therefore we represent this polynomial by Horner's method as a p -polynomial in $x - 8$.

$q=8$	1	-15	68	-119	87
		8	-56	96	-184
	1	-7	12	-23	-117
		8	8	16	
	1	1	20	17	
		8	72		
	1	9	92		
		8			
	1	17			

Hence $117q_1^4 - 137q_1^3 - 92q_1^2 - 17q_1 - 1 = 0$.

By means of arithmetic we see that for $q=2$ the last coefficient is still positive, but for $q=1$ it becomes negative, so q is in the interval $(1, 2)$. We have to make the Horner development for $q=1$.

$q=1$	117	-137	-92	-17	-1
		117	-20	-112	-120
	117	20	112	-120	
		117	14	14	
	117	137	-14	14	
		117	214		
	117	214	14		
		117			
	117	331			

In the same manner as it has been done for q we see that q_2 is situated in the interval $(1, 2)$.

$q=1$	130	144	-109	-231	-117
		130	274	75	-956
	130	4	104	-176	
		130	104	225	
	130	134	104		
		130			
	130	264			
		130			
	130	394			

q_2 is in the interval (2, 3)

$q = 2$	873	-223	-1013	-884	-130
		0	1046	66	-1166
		3	3	68	1
		11	2538	5142	
	773	1200	2571	4544	
		4	410		
	173	2013	1200		
		12			
	173	2013			

Presumably the reader has a new experience that q has to be chosen so that the sign of the last coefficient does not change, but that the procedure adopted for $q=1$ would alter the sign. $q=1$ will alter the sign of the second coefficient in the second row. Hence a scheme that the third coefficient will not change its sign, 6091 being too big. Hence the first coefficient will increase and therefore will not become negative. However for $q=1$, 6091 is outbalanced by more than 1000, and therefore $q=1$ is more than 1200, and the sign of the last coefficient would be altered. Hence $q=1$.

$q = 1$	1320	-4011	-201	-2701	13
		1304	201	11244	120000
	1120	760	1	100	120000
		201	11244	120000	
	1320	1201	2000	1	
		1304	11244		
	1320	11200	1		
		1304			
	1320	1201			

The next q will become = 1

P_1	=	8
P_2	=	9
P_3	=	17
P_4	=	43
P_5	=	180
P_6	=	242

Hence $\xi = (8, 1, 1, 2, 4, 1, \dots)$

Q_1	=	1
Q_2	=	3
Q_3	=	2
Q_4	=	5
Q_5	=	22
Q_6	=	27
Q_7	=	40

Now $\frac{1}{2} < \frac{1}{2}$ the error $\frac{1}{2} - \frac{1}{2} = 0$ is positive and $\leq \frac{1}{2}$, thus

$A_2 = \frac{1}{2}$ is the value of A when $A = \frac{1}{2}$ the second decant

only the third decant may be of 1. From this example the reader will see that Lagrange's method is sometimes very convenient for fraction calculation.

In the last subsections Horner's scheme has been used for the solution of equations with real coefficients by real roots. But this scheme can be applied as it has been stated in the beginning of this chapter to arbitrary fields. We will use it now to find out a theorem on complex numbers. Let b_0, b_1, \dots, b_n be the coefficients of a polynomial

$$P(x) = b_0 + b_1x + b_2x^2 + \dots + b_nx^n$$

$$= b_0 + b_1x + b_2x^2 + \dots + b_nx^n$$

$$= b_0 + b_1x + b_2x^2 + \dots + b_nx^n$$

Hence if a is a root, $\sum b_i a^i = 0$.

Let b_i be positive numbers and $a = re^{i\theta}$.

1) If $|a| < 1$ the equation cannot hold.

2) If $|a| = 1$, $a = \cos \theta + i \sin \theta$,

$\sum b_i a^i = \sum b_i (\cos i\theta + i \sin i\theta)$ hence $\sum b_i \cos i\theta = 0$ and $\sum b_i \sin i\theta = 0$ but in this case the last coefficient $a^n \sin \theta = \sum b_i \sin i\theta = 0$ and $b_n > 0$ is a contradiction.

If therefore $a = re^{i\theta}$, $r > 1$, $a^n \sin \theta = 0$.

$$a_0 < a_1 < \dots < 1, \quad (1)$$

then for every root a of this polynomial $|a| < 1$. Hence

$y = \frac{1}{x}$, the roots x of $\sum a_i x^i$ satisfy $|x| < 1$. This theorem is

known as

Littlewood's theorem. The complex roots of $\sum a_i x^i$ have absolute values < 1 if the coefficients satisfy (1).

§ 2. THE ROOTS OF REAL POLYNOMIALS

[2/1] In this section

$$a, b, c, d, e \quad (2.1)$$

will be with \bar{a} unless indicated by a bar over the symbol, and the same notation will be used for the conjugate of a complex number.

$$a = \alpha + i\beta, \quad \bar{a} = \alpha - i\beta \quad (2.2)$$

for α, β real numbers, α and β being ≥ 0 when a is not real.

Hence $a + \bar{a}$ is real; $a - \bar{a}$ is positive, $i - \bar{i} = 2i$, and if $a = 0$, $\alpha = \beta = 0$.

$$f(x) = a_0 + a_1x + \dots + a_{n-1}x^{n-1} + x^n \quad (2.3)$$

$$\text{can be represented by } f(x) = \prod_{i=1}^n (x - \alpha_i) \quad (2.4)$$

(see Part II [18.2]).

Theorem. If a is a root of (2.4) \bar{a} is also a root of (2.4).

1st Proof. Let K be the field of real numbers, i and $-i$ be the roots of $x^2 + 1$, then $K(i) = K + iK$ is the field of the complex numbers and there is an automorphism J of this field interchanging i with $-i$ and leaving the real numbers unaltered. $f(x)$ will not be altered by J , hence a will be transformed into a root of $f(x)$ but as a will be transformed to \bar{a} the theorem is true.

2nd Proof. If a is real, $a = \bar{a}$. If a is not real ($a = \alpha + i\beta$) $f(x)$ is a real polynomial and irreducible in the field of the real numbers. As $f(x)$ and $f(\bar{x})$ have a common root these polynomials have a common factor of positive degree. Hence $f(x)$ is divisible by $f(\bar{x})$ and \bar{a} is therefore a root of $f(x)$.

Corollary 1.

$$f(x) = (x - \alpha_1) \dots (x - \alpha_r) (x + i\beta_1)(x - i\beta_1) \dots (x + i\beta_k)(x - i\beta_k) \quad (2.5)$$

where $n = r + 2k$

Theorem 2.5 Every real polynomial has an odd number of real roots. In particular, if n is odd, then the number of real roots is at least one.

Corollary 2.6 If n is odd, there exists at least one real root.

$$\Delta = \prod_{1 \leq i < j \leq n} (a_i - a_j)^2 \quad (2.6)$$

is called the discriminant of $f(x) = \Delta_n(x)$. It is a symmetric polynomial in a_1, \dots, a_n with integral coefficients. It follows from Part II (10) that

$$\Delta = g(a_1, \dots, a_n)^2 \quad (2.7)$$

where g is a polynomial with integral coefficients. From (2.6) it follows (see Part II, (10/8)) that

$$\Delta = 0 \text{ if and only if } a_i = a_j \text{ for } i \neq j. \quad (2.8)$$

Let the n roots a_i be all different. As it has been proved in Part II (1-4),

$$\Delta = (-1)^{n(n-1)/2} \begin{vmatrix} a_1^{n-1} & a_1^{n-2} & \dots & a_1 \\ a_2^{n-1} & a_2^{n-2} & \dots & a_2 \\ \vdots & \vdots & \ddots & \vdots \\ a_n^{n-1} & a_n^{n-2} & \dots & a_n \end{vmatrix}^2 \quad (2.9)$$

To get Δ we have to interchange every number with its conjugate. From (2.9) it follows that the permutation σ which changes i to \bar{i} in the determinant (2.9)

Hence $\Delta = (-1)^k \Delta$ and therefore

$$(-1)^k \Delta = (-1)^k \Delta^2 = \Delta > 0$$

Hence the following theorem holds.

Theorem 2.7 Let $f(x)$ of degree n have n different roots. Then the discriminant of $f(x)$ is positive (negative) when the number of pairs of conjugate non-real roots is even (odd).

Corollary 2.8 A real polynomial of degree 3 has three real roots if and only if the discriminant is positive.

A real polynomial of degree 4 with positive discriminant has either four different real roots or two pairs of conjugate complex roots.

Exercise 2.9 Prove the preceding theorem without the help of (2.9).

[4.2] If $w(x)$ is a real-valued continuous and differentiable function. Hence

1. If $a < b$ and the signs of $w(a)$ and $w(b)$ are different then there is a root of $f(x)$ in the interval (a, b) .

2. If $a < b$ and $w(a) = w(b)$ then there is a root of $f(x)$ in the interval (a, b) .

3. If $f(x) = (x-a)^k g(x)$, $g(a) \neq 0$, $k > 0$, then

$$f'(x) = (x-a)^{k-1} g_1(x), \quad g_1(a) \neq 0$$

$$\text{For, } g_1 = k g(x) + (x-a) g'(x)$$

Then $f_1(x) = (x-a)^{k-1} g_1(x)$ is a function which has a root of order $k-1$ at $x=a$. Then $f_1(x)$ has a root of order n in each of the m different intervals $(a_1, a_2), (a_2, a_3), \dots, (a_m, a_{m+1})$ such that the multiplicity of the root is considered as n if f_1 does not change sign at a_i and then containing exactly n roots of $f(x)$ in the interval (a_i, a_{i+1}) . Every part of an interval in which $f(x)$ does not change sign is called a signum of f . The sign of f is different when x is a root of f of odd order. The sign is not different if x is a root of even order. Thus f_{k-1} is a nontrivial root of $f(x)$ of order k if k is an odd part of order q of $f(x)$.

The property holds for any analytic function with a finite number of roots and no singularities. If $f(x)$ is a polynomial with all coefficients of $f(x)$ are all positive all n signs of $f(x)$ becomes obviously positive for very positive values of x . Hence $f(x) \neq 0$ for any positive value of x . Thus

$f(x)$ is a positive root if there is no change of the signs in the sequence of coefficients of $f(x)$. So we are interested in the connection between the signum of $f(x)$ and the sign of the coefficients. The experience we get in solving all these problems is very helpful to us.

By developing $f(x)$ as a polynomial in $x-b$ we get

$$f(x) = f_1(x-b) = a_{0,b} + (x-b)a_{1,b} + \dots + a_{n,b}x^n \quad (2.10)$$

So for every real number b and the fixed $f(x)$ there belongs a set of $n+1$ real numbers $a_{0,b}, a_{1,b}, \dots, a_{n,b}$

such that $a_{0,b}, a_{1,b}, \dots, a_{n,b}$ independent of b (2.11)

$$a_{0,b} = f(b)$$

Hence $\gamma_{k,n} = 1$ if and only if $\gamma_{k,n}(b)$ is an even multiple of $\pi/2$ and $1/n < 1/k < n$.

$$\gamma_{k,n} = \frac{1}{k!} f^{(k)}(b). \quad (2.12)$$

The sequence

$$a_{1,n}, a_{2,n}, \dots, a_{k,n-1}, \dots, a_{n,n} \quad (2.13)$$

may be reduced by removing all even elements except $a_{1,n}$. If in this reduced sequence, $a_{k,n}$ has a sign different from the sign of the preceding element, it contributes $n-k+1$ to $C(b)$. We will denote the number of such changes

$$0 \leq C(b) \leq n \quad (2.14)$$

If the first and the last element of the sequence $a_{1,n}, \dots, a_{n,n}$ have the same sign, $a_{1,n}$ is an even number, if they have different signs $(-1)^{n-1} = -1$ from (2.11). It follows therefore that $C(b)$ is even if $a_{1,n} > 0$, where $a_{1,n} = f^{(1)}(b)$, and it is odd where $a_{1,n} < 0$.

If we strike out the 2^i th element which has not yet changed, $C(b)$ is different from the first, the number of changes will be unchanged, if we strike out an arbitrary element, the number of changes will not increase. In order to decrease $C(b)$, the function has to be a whole and if there is a whole, we will compare the changes in the rows

$$\begin{aligned} a_1 &= a_{1,1}, & a_2 &= a_{2,1}, \\ a_3 &= a_{3,1}, & a_4 &= a_{4,1}, \\ & \vdots & & \vdots \\ a_n &= a_{n,1}, & a_{n+1} &= a_{1,2}, \end{aligned}$$

where

$$a_{n+1} = a_{n,2}, \quad a_{n+2} = a_{2,2}, \quad a_{n+3} = a_{3,2}, \quad \dots$$

Let $a_1 > 0$. If $a_2 < 0$, there is a change in the 2^1 row, then $a_{n+1} < 0$ and it is of the same sign as a_1 . Hence the 2^1 row has the same number of changes as the sequence formed by these elements a_1 and the first row, which have the same sign as the corresponding elements of the second row. The number of changes of the 2^1 row is therefore less or equal to the number of the changes in the first. The same holds for every pair of subsequent rows in the following system which we get by Hensel's scheme

$$\begin{array}{cccccc} a_{n+1} & a_{n+2} & a_{n+3} & a_{n+4} & a_{n+5} & a_{n+6} \\ a_n & a_{n+1} & a_{n+2} & a_{n+3} & a_{n+4} & a_{n+5} \\ a_{n-1} & a_n & a_{n+1} & a_{n+2} & a_{n+3} & a_{n+4} \\ a_{n-2} & a_{n-1} & a_n & a_{n+1} & a_{n+2} & a_{n+3} \\ \vdots & \vdots & \vdots & \vdots & \vdots & \vdots \\ a_2 & a_3 & a_4 & a_5 & a_6 & a_7 \\ a_1 & a_2 & a_3 & a_4 & a_5 & a_6 \end{array}$$

But the last row is identical with

$$f_0(x) = f_0(x-1) + f_0(x-2) + \dots + f_0(x-n) + 0$$

Hence for $q > 0$, $C(f) \leq C(b)$ holds. (2.15)

Let $c = b + q > b$

$$f(x) = f_0(x-b) = f_0(|x-b|+q)$$

thus if f is in \mathcal{F} then the number of changes in f is not greater than the number of changes in f_0 , i.e.

$$C(c) \leq C(b) \quad (2.16)$$

As $f = x$ and $f_0 = x$ have the same number of changes we shall suppose with out any loss of generality that $c_1 > 0$. If $c_1 = c_{2n-1}$, and q is positive $c_{2n-1} > c_{2n-2} > c_{2n-3} > \dots > c_{2n-2}$ and if q is great enough $c_{2n-1} > c_{2n-2}$. In this manner we see that from a certain value of q the coefficients become all positive. Hence $C(c) = 0$ for $c > c$. Therefore

Theorem. C decreases steadily to 0

Let $b < \dots$ and let these values be not separated by a root of $f(x)$ then $c_1 = f(b)$ and $c_{2n-1} = f$ have the same sign. Hence C for $t = b + 2k$, where $k \geq 0$ is an integral number

C is therefore a decreasing function taking only integral values and the values at points which are not roots have even values. We have now to investigate the effect in the roots of $f(x)$.

Let c be a root of multiplicity m , and $x - c = y$, then

$$f(x) = a_0 y^m + a_1 y^{m-1} + a_2 y^{m-2} + \dots + a_{m-1} y + a_m = 0.$$

Let $y = x - c = t + x = t$

$$f(x) = f_0(t) = a_0 t^m + a_1 t^{m-1} + a_2 t^{m-2} + \dots + a_{m-1} t + a_m \quad (2.17)$$

Let $a_{i+m} y^m = a_{i,m} (t+x)^m = g(x)$

If $t > 0$ there is no change in the coefficients of $g(x)$

If $t < 0$, the coefficients have alternating signs and have therefore m changes. If t is small enough the last $m+1$ coefficients of $f_0(t)$ have the same signs as $g(x)$. If therefore c increases from a small negative to small positive values, the number of the changes in $a_{1,m}, \dots, a_{m,m}$ decreases by an

even value and the number of the changes n_1, n_2, \dots, n_m decrease by m .

Hence $C(b) \equiv C(a) + 2l + n_m$, where $l \geq 0$ is an integral number. By the help of these inequalities we get the following theorem.

Theorem. Let $f(x) = 0$, $f'(x) \neq 0$, $b < c$ and let r be the number of the roots of $f(x)$ in the interval (a, b) , every root being counted with its own multiplicity, then

$$C(b) \equiv C(c) + r + 2k, \quad (2.14)$$

where $k \geq 0$ is an integral number.

Applying the theorem (2.14) to $f(x) = x^n$ when $a = 0$ and $c = 1$ and that $C(c) \equiv 0$, we get as corollary

Descartes' rule. The number of the positive roots of $f(x) = 0$, root being counted with its own multiplicity, is equal to the number of the changes of sign of the coefficients of $f(x)$, r to a number less than or equal to an even number.

If we consider that in (2.12) a_{i+1} and $f^{(i)}(b)$ have the same sign, (2.14) can be expressed in the following manner:

Lemma 1. Let $f(x) \neq 0$, $f'(x) \neq 0$, then the number of the roots of $f(x)$ in the interval (b, c) , every root being counted with its own multiplicity, is equal to the difference of the changes of sign in the sets

$$f(b), f'(b), \dots, f^{(n)}(b) \text{ and}$$

$$f(c), f'(c), \dots, f^{(n)}(c) \text{ in an even number less than } n.$$

If we set $x = \frac{y+b}{y+1}$ and therefore $y = \frac{x-b}{1-x}$ then the positive roots of

$y(y+1)^n = 0$ correspond to the roots of $f(x)$ in the interval (b, c) . Therefore we are able to apply Descartes' rule to find out the number of the roots in this interval.

These formulas (2.14) give directly the exact number of the roots in an interval, but they are very useful for getting it even in more complicated cases.

We will go into further details of the example considered in (2.1).

$$f(x) = x^6 - 15x^5 + 68x^4 - 119x^3 + 67x^2.$$

As we stated before, the real roots are positive and situated in the interval $(1, 2)$. We can get this result on considering the change of sign. $f(1) = 4$ has no change and therefore no positive root. $f(2) = 8$ has no negative root. From the previous calculations for this example we get—on considering the signs only—

$$\begin{aligned} C(0) &= C(1) = 4 \\ C(2) &= 8 \\ C(8) &= 1 \end{aligned}$$

We know two roots, one in the interval $(1, 2)$ and another > 8 . The interval $(2, 8)$ contains each of these roots; there is no sign change. We showed that in the case of a change in sign, it corresponds to roots of $f(x)$. For this purpose, we try to approximate these expected roots by Horner's scheme and get by very simple calculations

$$\begin{aligned} C(3) &= 1 & C(2.6) &= 8 \\ C(2.65) &= 1 & C(2.64) &= 8 \end{aligned}$$

Hence the two roots can only be situated in the interval $2.64 < x < 2.65$ but we will prove that $f(x)$ is negative in this interval. We stated previously that

$$\begin{aligned} f(x) &= x^3 - 2x^2 + 1 = (x-1)(x^2 - x + 2) + 1 = (x-2)(x-3)(x-2) + 1 \quad (x > 0) \\ &= 2 - (x-1)(x-2)(x-3) + 1 = 3 - (x-1)(x-2)(x-3) \quad \text{Hence for } 2.64 < x < 2.65 \\ f(x) &= 2 - 1(x-1)(x-2)(x-3) < 0 \quad (x-1)(x-2)(x-3) > 1 \quad (2) \end{aligned}$$

Hence x has only the two roots expected in (1). The same result can also be obtained by calculating the discriminant and stating that it is negative.

[2/8] A method to get the exact number of the different roots of any n th degree has been given by Sturm. We suppose that $f(x)$ is a real function that therefore $f(x)$ has only single roots. There are no cases of pair roots, x is the real of $f(x)$ or y is a more accurate always obtained by the algorithm and if f should be of positive degree, $f(x)$ has to be replaced by $f(x) - f(x) - f(x)$, which has the same roots but on y assumptions.

The method uses a chain of Sturm's chain of polynomials

$$f(x) = p_0(x), p_1(x), \dots, p_n(x), \quad (2, 10)$$

and the number $C = 0$ of the changes of sign in

$$f_1(b), \dots, f_n(b).$$

The chain should be made in such a manner that $C(b)$ is a steadily decreasing function changing its value only at the roots of $f(x)$ and having at these points a value of the value ± 1 . Then

$$C(b) = C(a) - 1 \quad \text{for } b < a, \quad b \neq 0 \neq f \quad (2.2)$$

becomes the number of the roots of $f(x)$ in the interval b .

For this purpose we have to arrange the chain such that at each root of $f(x)$, one change will be lost and that in the roots of $f(x) > 0$ the number of the changes will be increased. In order to get a series of changes in the roots of $f(x)$, f' must take a sign of f' when it passes a root of $f(x)$ from the left to the right, i.e.

$$(1) \quad f_2(x) \text{ has the same sign, as } f'(x)$$

In order to avoid an increase in the value of C at roots of $f(x) < 0$

$$(2) \quad f_n(x) \text{ should have constant sign, and}$$

$$\text{for every root } b \text{ of } f(x) \quad f_{n-1}(b) \neq 0, \quad f_{n-2}(b) < 0.$$

So $f_{n-1}(x)$ will have the sign of exactly one of its neighbours before and after passing a root. The essence of these considerations is the following theorem

Sturm's theorem. If the chain (2.1) satisfies the conditions (1, 2), the number of the roots of $f(x)$ in any interval b_1, b_2 is given by $C(b_1) - C(b_2)$.

In order to get a chain of this kind we may use the algorithm to find the b.c.f.

$$f_2(x) = f'(x), \quad f_3(x) = q_1(x)f_1(x) - f_{1,2}(x) = f_{1,3}(x),$$

where the q_i are the p_i in the division of $f_1(x)$ by $f_2(x)$.

The first two conditions are obviously satisfied. As every common factor of $f_1(x)$ and $f_2(x)$ must be a common factor of $f_1(x) = f(x)$ and of $f_2(x) = f'(x)$ and as $f(x) < 0$ and $f'(x) > 0$ and $f_{1,2}(x)$ have no common root. If there are $f_{1,2}(b) = 0$, $f_{1,3}(b) = -f_{1,2}(b) \neq 0$. Hence

$$f_1(b)f_{1,3}(b) < 0.$$

By Sturm's method it is always possible to get the exact number of the different roots in any interval, but the practical calculation is sometimes very bothersome, so that it is often more convenient to use Budan-Fourier's theorem in connection with special considerations as we did it in the previous example. The reader may investigate the same example with Sturm's method as an exercise.

Remark. Sturm's chain (2, 19) can always be replaced by

(2, 20) $P_0(x), P_1(x), \dots, P_n(x)$ where c_1, c_2, \dots, c_n are positive constants

[2.4] Sturm's theorem with a biquadratic f and f' as polynomials

$$P_m(x) = \frac{1}{2^m} (x^2 - 1)^m \quad ; \quad m = 0, 1, 2, \dots \quad (2, 21)$$

D^m denoting the m -th derivative of the expression written in [] and D^0 is the function itself. If u and v are polynomials in x

$$D^m(uv) = \sum_{i=0}^m \binom{m}{i} D^i u \cdot D^{m-i} v \quad (2, 22)$$

$$D^m(x^2 - 1)^m = \binom{m}{0} (x^2 - 1)^m + \binom{m}{1} D(x^2 - 1)^{m-1} + \dots + \binom{m}{m} D^m(x^2 - 1)^0 = 2mx \cdot (x^2 - 1)^{m-1}$$

where \mathbf{I} in (2, 22) it follows for $m > 1$

$$D^m[(x^2 - 1)^m] = 2mx \cdot D^{m-1}[(x^2 - 1)^{m-1}] + \dots + m(m-1) D^{m-2}[(x^2 - 1)^{m-2}] \quad (2, 23)$$

On the other hand we get from (2, 22) for $m > 1$

$$\begin{aligned} 2D^m[(x^2 - 1)^m] &= 2D^m[(x^2 - 1)^{m-1}(x^2 - 1)] = 2D^{m-1}[(x^2 - 1)^{m-1}] \cdot D(x^2 - 1) \\ &\quad + 4x \cdot D^{m-1}[(x^2 - 1)^{m-1}] = 2m(m-1)D^{m-2}[(x^2 - 1)^{m-2}] \end{aligned} \quad (2, 24)$$

By subtraction of (2, 23) from (2, 24) and applying (2, 21) we get

$$mD^m[(x^2 - 1)^m] = x^2 D^{m-1}[(x^2 - 1)^{m-1}] + m x D^{m-1}[(x^2 - 1)^{m-1}] \quad (2, 25)$$

$$\text{As } D^0(x^2 - 1)^0 = 1 \quad x \neq \pm 1 \quad D^1(x^2 - 1)^0 = 0,$$

(2, 25) holds also for $x = \pm 1$ and therefore generally

From

$$D^{m-1}[(x^2 - 1)^{m-1}] = D^{m-1}[2mx(x^2 - 1)^{m-2}]$$

$$2mx D^{m-2}[(x^2 - 1)^{m-2}] + 2m^2 D^{m-2}[(x^2 - 1)^{m-2}]$$

we get

$$P'_m(x) = x P'_{m-1}(x) + m P_{m-1}(x) \quad (2, 26)$$

From (2, 25) and (2, 26) we eliminate P'_{m-1} and get

$$(x^2 - 1)P'_m(x) = mx P_m(x) - m P_{m-1}(x) \quad (2, 27)$$



On replacing m by $m+1$ in (2, 25) we get

$$(m+1)P_{m+1}(x) = (x^2-1)P'_m(x) + (m+1)xP_m(x). \quad (2, 25')$$

From (2, 27) and (2, 25') we eliminate $P'_m(x)$ and we get

$$(m+1)P_{m+1}(x) = (2m+1)xP_m(x) - mP_{m-1}(x). \quad (2, 26)$$

We consider the sequence

$$P_n(x), P_{n-1}(x), \dots, P_1(x), P_0(x) = 1 \quad (2, 27)$$

in the interval

$$-1 \leq x \leq +1.$$

From (2, 27) it follows that if $P_n(x) = 0$, $P_{n-1}(x)$ has the same sign as $P_{n-2}(x)$. If $P_n(x) = 0$ and $P_{n-1}(x) = 0$ we have a common root. This root has also as a root $P_{n-2}(x)$ as we see from (2, 26) and (2, 27). Hence for $P_n(x) = 0$, $P_{n-1}(x) \neq 0$ and $P_{n-2}(x) \neq 0$ it follows from (2, 26) that $P_{n-1}(x)P_{n-2}(x) < 0$ at every root of $P_n(x)$.

As $P_n(x)$ and $P_{n-1}(x)$ have no common root in the interval $-1 \leq x \leq +1$ the $P_n(x)$ has no common root with $P_{n-1}(x)$ in the interval $-1 \leq x \leq +1$. If $P_n(x) = 0$ in the interval $-1 \leq x \leq +1$ for every n .

From (2, 27) it follows that

$$P_n(1) = P_{n-1}(1)$$

and

$$P_n(-1) = -P_{n-1}(-1).$$

As $P_0(x) = 1$, it follows that

$$P_n(1) = 1$$

and

$$P_n(-1) = (-1)^n.$$

The number of changes of sign (2, 28) is therefore n if n is odd and $n-1$ if n is even. Hence there are n different roots in the interval $-1 \leq x \leq +1$ if n is odd and $n-1$ if n is even. Hence the roots of Legendre polynomials are all situated in the interval $-1 \leq x \leq +1$ and are simple roots.

To find out systematically the real roots of a polynomial

[2-1]

$$f(x) = a_0 + a_1x + \dots + a_nx^n$$

with real coefficients we have at first to find an interval containing all these roots.

Let

$$f \approx 1 + \left| \frac{a_n}{a_0} + \frac{a_1}{a_0} + \frac{a_2}{a_0} \right|$$

As a_0 and $1 + f$ have the same sign as a_n , and therefore f has the same sign as $a_0 - 1 + f \approx f$. Hence the roots of $f(x)$ are situated in the interval $(-f, +f)$.

When an interval containing all the real roots of the polynomial has been found, we can approximate its length as much as we wish. Any roots are contained in the subintervals. The subintervals containing roots should be subdivided again, and the procedure must be repeated, till an interval x has been found for every root b such that $x < b < x + \epsilon$ and ϵ is as small an approximation required for the given problem. As the root has to be represented by a decimal fraction, the first subdivision will usually be made by the integral value of x , the second subdivision by numbers which will be more than after $m + 1$ subdivisions the root will be determined up to the m^{th} decimal.

[2.6]

This method is suitable for points which are to be used by a calculator. By the practical reckoning, to which it corresponds, a whole subinterval may contain roots, and by subdividing the polynomial only for the endpoints of these intervals. If an interval containing only one simple root, the approximation in m steps can be obtained very quickly by Horner's scheme.

Example 1. If a and b have different signs, the graph of the function $f(x)$ may be replaced by the straight line connecting the points with abscissas a and b . This line intersects the x -axis in $x = \frac{a}{b-a} \cdot (b-a) = \frac{ab}{b-a}$. This value may be considered as a first approximation of the root. For the example of (1)

$$x^3 - 1 - y = f(x) = f(y) = y^3 - 11y^2 + 20y - 2$$

$$g(0) = 2, g(1) = -3.$$

Hence we get by the *Example 1* as a first approximation $y_1 = 1/2$. We will try to improve this approximation by

$$24x^2 - y^4 = 11y^3 + 20y^2 + 2 \quad (2.30)$$

Hence

$$y_2 \approx 1/4.$$

This approximation is better, but it is not good. As we calculated in §1

$$x - 1 \approx 0.0035324$$

Of course the graph of the polynomial is in that interval very different from a straight line. Now

$$\begin{aligned} y(0) &= 2 & y(1) &= 1 \\ y'(0) &= -24 & y' &= \end{aligned}$$

The graph is therefore concave down in the interval and we must obviously be near the point $x - 1 \approx 0$.

Newton's method. The graph is now approximated by the tangent to the term of higher degree in Horner's scheme have to be omitted. This method gives good results when the distance from the root is small. The term of $(x - 1)^2$ degree has a constant of high absolute value and the absolute values of the coefficients of the higher terms are not too big. This is the reason why it is used in the above example, and gives the approximation $2 - 24\sqrt{x - 1}$. We can improve this solution by using the equation y_1 with $y_1 = 0.1$ then y_2 becomes < 0.005 .

As for $y_1 = 0, \quad y_2 > 0.002$

and for $y_1 = 0.1, \quad y_2 < 0.003$ hold,

and as y_2 is a continuous function of y_1 there must exist in the interval $(0.002, 0.003)$ a value for which $y_1 = y_2$, and that is a root. So Newton's method is very useful in certain cases, but if the interval is big or the tangent makes only a small angle with the axis the method cannot be used.

The suitable choice of the methods should be learned by practice. In this section sometimes reference has been made to the graph of a polynomial. So the reader may ask if graphical methods may not be helpful to get the roots of a polynomial. Of course methods of this kind exist and are very helpful to get a convenient first approximation for the roots, but in applying these methods only those readers may succeed who are familiar with the theory and practice of mathematical drawing.*

Let x be a fixed real number $\neq 0$ and let b and c be two different and consecutive roots of (1) then $f(x)$ has a constant sign in the interval

[2.7]

* A very useful graphical method is e.g. the rectangle method. For reference see Fiebertsch-Kauer, Vorlesungen über Algebra, pp. 134-142.

Let $h_{j+1}(x) = x^j + \dots + a_{j+1}x + a_j$. As the theorem holds for $m = j$, the number of the roots of $g_j(x) = \sum_{i=0}^j a_i f^{(i)}(x)$ is greater than or equal to the number of the roots of h_{j+1} . In the formula $f^{(i)}(x)$ means $f^{(i)}(x)$.

$$g_j(x) = \sum_{i=0}^j a_i f^{(i)}(x).$$

If we replace in $h(x) = a_0 x^n + \dots + a_1 x + a_0$ the place of f by the corresponding derivatives of $f(x)$, we get therefore

$$g(x) = g_1(x) - a_1 g_0(x).$$

From the preceding theorem it follows that the number of the roots of $g(x)$ is not less than the number of the roots of $g_1(x)$ and therefore not less than the number of the roots of h . Hence the theorem is proved.

1.5. GRAPPE'S METHOD

By Grappe's method all the roots can be calculated with a precision [3.1] measured at the point t . The method consists of the following steps: every quantity which has to be calculated is written as a function of t , and we estimate exactly the error made by substituting t for x . Hence the roots have to be estimated by putting t for x .

$$\text{Let } h_1 > h_2 > \dots > h_n \quad (3.1)$$

be the roots of the polynomial $a_0 x^n + a_1 x^{n-1} + \dots + a_n$, then

$$\frac{-a_1}{a_0} = h_1 + \frac{h_2}{b_1}, \quad \frac{h_2}{b_1} = h_2 + \frac{h_3}{b_2},$$

$$\frac{-a_2}{a_0} = \frac{-a_2}{a_0} - \frac{a_1}{a_0},$$

$$t_1 t_2 \left(1 + \frac{h_2}{b_1} + \frac{h_3}{b_1 b_2} + \frac{h_4}{b_1 b_2 b_3} + \dots + \frac{h_n}{b_1 b_2 b_3 \dots b_{n-1}} \right) = \frac{-a_2}{a_0} - \frac{a_1}{a_0}.$$

$$t_2 = 1 + t_1 t_2$$

...

$$\frac{t_{n-1}}{t_{n-2}} = b_{n-1} + \frac{h_n}{t_{n-2} b_{n-1}}$$

$$\frac{-a_n}{a_0} = (-1)^n \frac{a_n}{a_0} \left(\frac{-t_1}{b_1} - \frac{a_1}{a_0} - \frac{-t_2}{b_2} - \dots - \frac{-t_{n-1}}{b_{n-1}} \right)$$

$$\frac{t_1}{b_1} + \frac{h_2}{b_1 b_2} + \dots + \frac{h_n}{b_1 b_2 b_3 \dots b_{n-1}} = b_n (1 + \dots)$$

If $b_i - b_{i-1}$ is very great (for $i = 1, \dots, n$) the numbers c_i can be omitted. In this case we get the approximation

$$b_i \wedge_{n-1}^{\infty} \quad \text{for } i = 1, \dots, n. \quad (B, 2)$$

The general \wedge is not a composition of \wedge and \vee and for a rational expression $a_0 + a_1 x + \dots + a_n x^n \wedge_{n-1}^{\infty}$ we have to find out a polynomial $f(x)$ such that $b_i = b_i^{(m)}$ (the i th root of this polynomial is denoted by b_i) for $i = 1, \dots, n$. It is not a priori clear that such a polynomial $f(x)$ exists. But it is easy to find out a polynomial $f(x)$ such that the spaces \wedge terminate ($f(x)$ and by repeating this construction we can subsequently polynomialize with them to

$$f(x) \wedge b_1 \wedge b_2 \wedge \dots \wedge b_n x^0$$

Let $f(x) = x^n + a_1 x^{n-1} + \dots + a_n$ have the roots $b_1^{(m)}, \dots, b_n^{(m)}$ and we can choose m great enough that $b_i^{(m)} - b_i$ can be omitted then $b_i = \wedge_{m-1}^{\infty} b_i^{(m)}$.

The polynomial $f(x)$ is polynomial with roots $b_1^{(m)}, \dots, b_n^{(m)}$, therefore we can approximate the spaces \wedge by \wedge and \vee approximately the spaces \wedge are completed and closed. Hence we have to repeat the construction of $f(x)$ and \wedge and \vee and so on that after further repetition \wedge and \vee are complete and so on the spaces of the \wedge and \vee of the previous process. Let $p(x)$ be a polynomial whose roots are the square roots of b_1, \dots, b_n we can write

$$\begin{aligned} p(x) &= (x^2 - b_1)(x^2 - b_2) \dots (x^2 - b_n) = (x^2 - b_1)(x^2 - b_2) \dots (x^2 - b_n) \\ &= f_p(x^2) \end{aligned}$$

The coefficients of $f_p(x)$ are calculated by the following arithmetic

(1)	a_0	a_1	a_2	\dots	a_n
	a_0	$-a_1$	a_2	\dots	$\pm a_n$
(2)	a_0^2	$-a_1^2$	a_2^2	\dots	$\pm a_n^2$
		$+ 2a_0 a_1$	$+ 2a_1 a_2$		
			$+ 2a_2 a_3$		

As in the first pair of lines corresponding numbers differ only by the sign it is usual to write only the signs in the 2nd row. The numbers increase very quickly, therefore it is convenient to omit the last figures denoting the decimals very easily. For this purpose we shall use the notation

$$8^{\circ}456181 \quad \text{for} \quad 8.456181 \cdot 10^8.$$

To extract the roots at the end of the calculation we need logarithms. It is therefore useless to calculate more decimals than the tables of logarithms contain.

Example.

$$x^3 - 10x^2 + 16x - 2 = 0$$

$$\begin{array}{rcl} (1) & 1 & - 1^{\circ}0 & 1^{\circ}6 & - 2 \\ & + & + & + & + \end{array}$$

$$\begin{array}{rcl} & 1 & - 1^{\circ}0 & 2^{\circ}4 & - 4 \\ & & + 0.32 & - 0.4 & \end{array}$$

$$\begin{array}{rcl} (2) & 1 & - 0^{\circ}8 & 2^{\circ}10 & - 4 \\ & + & + & + & + \end{array}$$

$$\begin{array}{rcl} & 1 & - 4^{\circ}024 & 4^{\circ}0350 & - 16 \\ & & + 0.432 & - 0.0344 & \end{array}$$

$$\begin{array}{rcl} (4) & 1 & - 4^{\circ}152 & 4^{\circ}6112 & - 16 \\ & + & + & + & + \end{array}$$

$$\begin{array}{rcl} & 1 & - 1^{\circ}75729 & 9^{\circ}12031 & - 256 \\ & & + 0.0032 & + 0.0018 & \end{array}$$

$$\begin{array}{rcl} (6) & 1 & - 1^{\circ}74797 & 9^{\circ}12618 & - 256 \end{array}$$

In the next step the coefficients will become the squares of the preceding coefficients, and in no case the error will have influence on the first 4 figures. Therefore we stop the procedure and calculate now the roots by the help of logarithms.

log. of the coefficients	log. x^2	log. $ x $	x
0	7.24254	0.90332	8.0412
7.24254	2.04506	0.20833	1.6228
0.72260	1.04404—8	0.13609—1	0.1365
2.40224			
		0.30308	10.0000
		= log 2	

The sign of the roots cannot be determined by Graeffe's method; we have to arrange a special investigation for the signs in every case. In this example the coefficients have alternating signs, hence the coefficients of $f(-x)$ are all positive. The real roots of $f(-x)$ are therefore negative, hence the roots of $f(x)$ are all positive.

For $x = 10, 100$ we form the elementary symmetric functions of the approximate roots, and we get

$$\begin{array}{ccccc} x_1 = 10, & x_2 = 15.0008 & x_3 = 2 \\ \text{for} & 10 & 100 & 2 \end{array}$$

[5.2] If a real polynomial has complex roots of the same absolute value, and these have therefore the same absolute value. Graeffe's method has therefore to be modified in this case. An example will give valuable hints for necessary modifications.

Example. $x^6 - 11x^3 + 20x - 24x + 2$.

We know from § 1 that this polynomial has two real roots $b_1 = 7.5926$ and $b_2 = 0.00324$ and two complex roots.

The calculation by Graeffe's method is given* on the next page.

If the procedure be repeated, the two first and the two last coefficients will become the squares of the corresponding coefficients of the line (6), but the third coefficient will depend also on the second and the fourth. We cannot expect that further repetition of the procedure will make the third coefficient independent of his neighbours, as two roots of the polynomial have an equal absolute value. If b_1 is greater than the absolute value of

* As the sign in the 3rd line is a + sign + we omit these signs for abbreviation.

	1	- 11	29	- 24	2
(1)	1	- 121	841	- 576	4
		68	- 528	+ 120	
			+ 4		
(2)	1	- 60	812	- 400	4
	1	- 3601	100180	- 211200	16
		+ 634	- 57000	+ 2480	
			4		
(4)	1	- 8835	42537	208064	16
	1	- 1711224	1790000	- 4108767	256
		+ 851	- 141526	0	
			6		
(8)	1	- 1713571	1790113	- 4108567	256

the complex roots, then

$$\frac{-b_1}{2a_1} = h_1 = \frac{(1 + 2b_2^m + b_3^m)}{b_1^m} \quad \text{with } m \text{ for a suitable } m.$$

A rough mental calculation shows that $b_1^2 \approx 60$, $b_1^3 \approx 1^2 2$.

The same consideration made for $f, \frac{1}{x}$, shows that if $b_4 < b_2$

$\frac{-b_4}{2a_4} \approx b_4^m$ holds. Hence the complex roots are only dependent on the

8 mod 8 coefficients. In order to get the law of dependence we shall generalize the considerations.

Let B and ϵ be two intervals so that every number of C is very small in comparison to the numbers of B and let

$$f(x) = a_p + a_{p-1}x + \dots + a_0x^p \quad n = p + 1$$

have two sets of roots

b_1, b_2, \dots, b_r , whose absolute values belong to B and

c_1, c_2, \dots, c_s , whose absolute values belong to C .

Let $n \rightarrow \infty$ then $\frac{1}{n} \log \frac{1}{n}$ becomes approximately equal to the n^{th} symmetric fundamental function of b_1, \dots, b_r and

$$\frac{1}{n} \log \frac{1}{n} = \frac{d_r}{d_r} + \left(\frac{d_r}{d_r} \right) s_r^{\frac{1}{n}},$$

where s_r is a suitably chosen mean value of the roots of the second set, and therefore a number of C . Let y be a number of B then

$$\frac{1}{n} \log \frac{1}{n} = \frac{d_r}{d_r} + \left(\frac{d_r}{d_r} \right) y^{\frac{1}{n}} + \left(\frac{d_r}{d_r} \right) y^{\frac{1}{n}} + \dots + \left(\frac{d_r}{d_r} \right) y^{\frac{1}{n}},$$

$$\sum_{i=1}^r \frac{1}{n} \log \frac{1}{n} = \frac{1}{n} \log \frac{1}{n} = \frac{1}{n} f_1(y).$$

Let $y = 1$ and n be one of the roots b_i , then $y = 1$, and we can therefore approximate the last set of the roots of $f(x)$ by the roots of $f_1(x)$. But as a common factor of the coefficients has no importance, the roots b_i are approximated by the roots of

$$x^n + a_{n-1}x^{n-1} + \dots + a_0 = 0.$$

Let $x = \frac{1}{y}$, then $x^n + a_{n-1}x^{n-1} + \dots + a_0 = 0$

has the roots $\frac{1}{y_1} = b_1, \dots, \frac{1}{y_r} = b_r$

$$\text{and } \frac{1}{b_1} = c_1, \dots, \frac{1}{b_s} = c_s,$$

the absolute values of b_i belong to an interval B , the absolute values of c_i belong to C and every number of C is small in comparison to B . Hence the roots b_i can be approximated by the roots of $x^n + a_{n-1}x^{n-1} + \dots + a_0 = 0$, and therefore the roots b_i of $f(x)$ can be approximated by the roots of

$$a_0x^n + a_{n-1}x^{n-1} + \dots + a_0 = 0.$$

So the polynomial $f(x)$ has to be split up in two polynomials, the first is defined by the $r + 1$ upper terms and leads to the upper class of roots,

the second one is defined by the $s + 1$ lower terms, and so on to the lower class of roots.

The two classes may also be divided into sub-classes etc. Finally we get classes

$$b_{1,1}, \dots, b_{1,r_1}, \quad b_{2,1}, \dots, b_{2,r_2}, \quad \dots, \quad b_{s,1}, \dots, b_{s,r_s}$$

each root being small in comparison with the roots of the preceding classes, and to each class corresponds a polynomial which can be cut out from $f(x)$. The ratio of the absolute value of the roots increases when we replace these roots by higher powers of them therefore we get finally by Grabbe's method k polynomials each of them having only roots with the same absolute value. In the previous examples these polynomials are

$$x^4 - 1'10371 - (1 + 171)x^2 + 2^{10}413x + 4^{10}3717, \quad 4^{10}3717x^2 - 2^{10}$$

From these polynomials we get the roots of $f(x)$

$$8 \log |b_1| = 7.01286, \quad \log |b_1| = 0.87660, \quad |b_1| = 7.02$$

$$8 \log |b_2| = 7.76339 - 10, \quad \log |b_2| = 1.96674 - 2, \quad |b_2| = 0.023332$$

$$16 \log |b_3| = \log 4^{10}3717 - \log 1'10371$$

$$= 0.79767, \quad \log |b_3| = 0.22479, \quad |b_3| = 1.07779$$

$$\log \cos 8\phi = \log 1^{10}413 - 8 \log |b_2| = -\log 2 - \log 110371 = 9.46263 - 10$$

$$8\phi = +73^{\circ}31' + k \cdot 360^{\circ}$$

$$\phi = \pm 9^{\circ}11'23'' + k \cdot 45^{\circ}$$

To finish this calculation we have to fix the signs of the real roots and to determine the integral number k . As the signs of the coefficients are alternating there is no negative root. Hence $b_1 = 7.02$, $b_2 = 0.023332$. These numbers correspond to the results obtained in § 1 by Horner's method and by Lagrange's method.

As $b_1 + b_2 + b_3 + b_4 = 11 - 2 \cos \phi = 15$. But as $2c = 3 \cos \phi$, ϕ must be a very small angle. Hence $k = 0$. So we get

$$b_1 = 1'0367 + i0'26803$$

$$b_2 = 1'0367 - i0'26803$$

$$\text{For checking } \log b_1 + \log b_2 + 2 \log 2 = 0.30104$$

$$\text{for } \log 2 = 0.30103$$

$$b_1 + b_2 + b_3 + b_4 = 10.0000$$

$$\text{for } 11$$

If we replace in this last verification the value for b_1 by the more exact value obtained from $\log b_1 = 7.2525$ we will get

$b_1 + b_2 + b_3 + b_4 = 10.0001$. The result can be corrected by further calculation. As we see from the results of § 1 and from the checking given here b_1, b_2 and c are very exact. The correction is therefore expected to concern mainly the angle ϕ whose true value may be a little smaller. As ϕ itself is a small angle this correction will materially affect $\sin \phi$. Hence the imaginary parts of b_2 and b_3 are true up to the second decimal only.

If a polynomial with roots of equal absolute value has a degree > 2 , either it has multiple roots or it has non-conjugate roots. The multiple roots will be removed, when we divide by the h.c.f. of the polynomial and its derivative. Non-conjugate roots of equal absolute value can be cleared away by Horner's scheme, viz. if $|x| = |x'|$ and x is different from x' and \bar{x} , then $|x - \bar{x}| \neq |x' - \bar{x}|$.

Hence the real and the complex roots of (1) can be found out by a combination of Graeffe's method and Horner's scheme in every case. The results should be verified and it is possible to minimise the error by the methods given in § 1.

§ 4. ROOTS OF COMPLEX POLYNOMIALS.

Let $\phi(x)$ be a polynomial with complex coefficients,

$$\phi(x) = a_0 + a_1 x + \dots + a_n x^n$$

$$\bar{\phi}(x) = \bar{a}_0 + \bar{a}_1 x + \dots + \bar{a}_n x^n,$$

$$\phi(x), \bar{\phi}(x) = f_1^2(x), \quad \phi(x) = f_1(x) \phi_1(x), \quad \bar{\phi}(x) = f_1(x) \bar{\phi}_1(x)$$

$$\phi_1(x) \bar{\phi}_1(x) = f_2(x),$$

then $f_1(x)$ and $f_2(x)$ are real polynomials. On applying Graeffe's method to these polynomials we get the roots of $\phi(x)$, but out of two conjugate

roots of $f_2(x) = na$ is a root of $\phi(x)$, the other is a root of $\bar{\phi}(x)$, and we have therefore to make a verification finally

Let $|x| \geq \sum_{i=0}^n \frac{1}{a_i} = t$, then

$$\frac{1}{a_0} |\phi(x)| \geq |x|^n - \sum_{i=1}^n \frac{1}{a_i} |x|^{n-i} \geq |x|^{n-1} (|x| - 1) = |x|^{n-1} > 0$$

hence $\phi(t) \neq 0$, and therefore the absolute value of the roots of $\phi(x)$ is $< t$.

Another limit for the roots can be found by Hakeya's theorem.

The roots of $\phi(x)$ are also roots of the real polynomial $f_1(x) = f_2(x) = f(x)$ where $x = z + a$ and the real number a can be chosen in such a manner that $f(x) = a_0x^n + \dots + a_1x + a_0$ has positive coefficients only.

For the polynomials with positive coefficients the following theorem holds

Theorem Let the coefficients of $f(x) = a_0 + a_1x + \dots + a_nx^n$

be positive and $0 < p < \frac{1}{a_0}, 1 < q$ for $k = 1, \dots, n$, then the roots of $f(x)$ have to satisfy the condition

$$p < |x| < q.$$

Proof Let $x = qy$, $f(x) = f(y) = \sum b_k y^k$ then $b_0 = q^0 a_0$,

Hence $b_{k+1} = b_k < 1$. From Hakeya's theorem it follows therefore, for the roots that $|y| < 1$, and $|x| < q$.

The roots of $F(s) = a_0s^n + \dots + a_{n-1}s + a_0$ are reciprocal to the roots of $f(x)$. As $\frac{1}{a_0} < \frac{1}{p}$ holds, it follows from the first part

of the proof that the roots of f have to satisfy

$$|x| < \frac{1}{p} \quad \text{Hence } |x| = \left| \frac{1}{x} \right| > p \text{ holds for the roots of } f(x)$$

An interesting connection between the roots of $\phi(x)$ and its derivative $\phi'(x)$ is given by the

Theorem 5.11. Every convex polygon including all the roots of $\phi(x)$ contains every root of $\phi'(x)$.

Proof. Without any loss of generality we can suppose that ϕ and ϕ' have no common root. Let γ be an arbitrary root of ϕ and β_1, \dots, β_n be the roots of ϕ' , then

$$\frac{\phi'(x)}{\phi(x)} = \sum_{i=1}^n \frac{1}{x - \beta_i}, \text{ hence } 0 = \frac{\phi'(\gamma)}{\phi(\gamma)} = \sum_{i=1}^n \frac{1}{\gamma - \beta_i} \text{ and therefore}$$

$$0 = \sum_{i=1}^n \frac{1}{\gamma - \beta_i} = \sum_{i=1}^n \frac{\gamma - \beta_i}{|\gamma - \beta_i|^2} = \sum_{i=1}^n (\gamma - \beta_i) \cdot b_i, \text{ where } b_i \text{ is positive}$$

We consider the geometrical representation of the complex numbers in the plane. $(\gamma - \beta_i) \cdot b_i$ are vectors starting from γ and directed to β_i , $i = 1, \dots, n$. As the sum is equal to 0, every component of this sum is equal to 0. Let G be an arbitrary straight line passing through γ . The components of $(\gamma - \beta_i) \cdot b_i$ orthogonal to G form a sum equal to zero, hence either the components are all equal to zero or there are components with different sign. In the 1st case the points β_i are all situated on G ; in the 2nd case, there are roots of ϕ on both sides of G . In no case there are roots of ϕ on one side of G only. Let now P be a convex polygon including all the roots of ϕ . If γ is outside of P we can draw a straight line G not intersecting P through γ . Hence P and therefore all the roots of ϕ are situated on the same side of G . Hence γ is not a root of ϕ' .

Let P_n be the smallest convex polygon including the roots of ϕ . (The reader may prove that such a polygon exists and is unique). P_1 the corresponding polygon defined by ϕ' , P_2 the smallest polygon containing the roots of ϕ' . The polygons with higher indices are included in the preceding one. P_n degenerates to the point $\frac{1}{n} \sum_{i=1}^n \alpha_i$, $\alpha_i = \frac{1}{n} \sum_{i=1}^n \alpha_i$. This point is the centre of gravity of the roots of ϕ and for the same reason it is the centre of gravity of the roots of ϕ' and of the roots of each derivate.

§ 5. INTERPOLATION

[5/1] Let

$$\beta_1, \dots, \beta_{n+1} \tag{5, 1}$$

be $n+1$ different elements of an arbitrary field K , and let

$$\lambda_1, \dots, \lambda_{n+1} \tag{5, 2}$$

be $n+1$ arbitrary elements of K .

We want to find out a polynomial f of $K[x]$ such that

$$f(x_i) = \lambda_i \quad \text{for } i = 1, \dots, n+1 \quad \text{and degree } f \leq n.$$

Let $f(x) = a_0 + a_1x + \dots + a_nx^n$. The polynomial has the proposed properties if and only if its coefficients satisfy

$$\sum_{j=0}^n a_j \beta_i^j = \lambda_i.$$

The determinant of this system of $(n+1)$ linear equations (see Part II [10, 11]) is equal to $\prod_{i < j} (x_i - x_j)$ and is $\neq 0$ as the $n+1$ elements x_i are supposed to be different. Hence the problem has one and only one solution. This solution can be obtained by the methods explained in Part I, but it is easier to get it from special cases.

Let $f_0(x)$ be the solution if $\lambda_0 = 0, \lambda_i = 1$, then $f = \sum_{i=1}^{n+1} \lambda_i f_i(x)$

is the solution for arbitrary λ -elements. Let $f_0(x) = \frac{f(x)}{\prod_{i=1}^{n+1} (x - x_i)}$

where $f(x) = \prod_{i=1}^{n+1} (x - x_i)$. So we get the following formula for interpolation

$$f(x) = \sum_{i=1}^{n+1} \lambda_i \frac{f(x)}{(x - x_i) f'(x_i)} \quad (3)$$

By Lagrange's formula the problem of interpolation has been solved in the most complete and general manner, but the formula is not convenient for practical calculation. It is easier to calculate the coefficients of the product representation of $f(x)$. [5/]

$$f(x) = \gamma_0 + \gamma_1(x - \beta_1) + \gamma_2(x - \beta_1)(x - \beta_2) + \dots + \gamma_n(x - \beta_1)\dots(x - \beta_n) \quad (4)$$

Here is $\gamma_0 = f(\beta_1) = \lambda_1$, $\gamma_1 = \frac{\lambda_2 - \lambda_1}{\beta_2 - \beta_1}$ and we may successively calculate the coefficients γ_k . It is convenient to arrange this calculation in the following manner.

Let $f_k(x)$ be defined by $f_k(x) = f(x)$, and for $k=1, \dots, n$

$$f_k(x) = \frac{f_{k-1}(x) - f_{k-1}(\beta_k)}{x - \beta_k},$$

then $f_k(x) = \gamma_0 + \gamma_{k-1}(x - \beta_{k-1}) + \dots + \gamma_n(x - \beta_1)\dots(x - \beta_n)$

Hence $f_1(\beta_{n+1}) = \gamma_1$. We have therefore to calculate the values

$$\{k, m\} = f_k(\beta_m) \text{ for } k=0, \dots, m, \quad k < m \leq n+1$$

by $\{k, m\} = [\{k-1, m\} - \{k-1, k\}] \cdot \beta_m - \beta_k$

and $\{0, 0\} = \lambda_0$. We calculate the values columnwise in the following scheme

	$\{0, m\}$	$\{1, m\}$	$\{2, m\}$...	$\{n, m\}$
$\{k, 1\}$	λ_1				
$\{k, 2\}$	λ_2	$\frac{\lambda_2 - \lambda_1}{\beta_2 - \beta_1}$			
$\{k, 3\}$	λ_3	$\frac{\lambda_3 - \lambda_1}{\beta_3 - \beta_1}$	$\frac{\{1, 3\} - \{1, 2\}}{\beta_3 - \beta_2}$		

$$\{k, n+1\} \lambda_{n+1} = \frac{\lambda_{n+1} - \lambda_1}{\beta_{n+1} - \beta_1} \cdot \{1, n+1\} - \frac{\{1, n+1\} - \{1, 2\}}{\beta_{n+1} - \beta_2} \cdot \{2, n+1\} - \dots - \frac{\{n-1, n+1\} - \{n-1, n\}}{\beta_{n+1} - \beta_n} \cdot \{n, n+1\}$$

The first elements of the different columns of this scheme form the set $\gamma_0, \gamma_1, \dots, \gamma_n$ of the coefficients of $x = 4$. This scheme is easier for calculation than Lagrange's formula.

[5/8] The reckoning can further be simplified if the elements $\beta_1, \dots, \beta_{n+1}$ are equidistant, i.e. if

$$\beta_{i+1} - \beta_i = \Delta_i$$

for every i ; then

$$\begin{aligned} \Delta_k \cdot \{k, m\} &= [\{k-1, m\} - \{k-1, k\}] \cdot (m-k) \\ &= \frac{1}{m-k} \sum_{i=k}^{m-1} \Delta_{i+1} \cdot (i-k+1), \end{aligned}$$

where $\Delta_{i+1} = \{i+1, i+1\} - \{i+1, i\}$ is the difference of two consecutive elements in the preceding column. So $\Delta \cdot \{k, m\}$ is the mean-value of the differences of consecutive elements of the rows m to k in the column $k-1$.

We will now transform the scheme for these cases in such a manner that we have to calculate the differences of consecutive elements only and not the mean-values.

For this purpose we introduce the notations usual in the calculus of differences.

Let $\Delta_x : u = x - \beta_1$, then $\Delta_x(u - k - 1) = x - \beta_1$.

Let $F(u) = f(x) = f(\Delta_x u + \beta_1) = \gamma_0 + \gamma_1 u \Delta_x + \gamma_2 u(u-1) \Delta_x^2$
 $+ \dots + \gamma_n u(u-1) \dots (u-n+1) \Delta_x^n$

and let $\Delta f(x) = f(x + \Delta_x) - f(x) = f(u+1) - f(u)$ then

$$\Delta f(x) = \Delta_x [\gamma_1 + 2\gamma_2 u \Delta_x + \dots + n \gamma_n u(u-1) \dots (u-n+2) \Delta_x^{n-1}]$$

$$\text{etc. } (u+1)u(u-1) \dots (u-k+1) - u(u-1) \dots (u-k) = (k+1)u(u-1) \dots (u-k+1).$$

Let $\Delta(\Delta f(x)) = \Delta^2 f(x)$, \dots , $\Delta(\Delta^{n-1} f(x)) = \Delta^n f(x)$,

then we get by repetition of the procedure

$$\Delta^2 f(x) = \Delta_x^2 [2\gamma_2 + 2\gamma_3 u \Delta_x + \dots + n(n-1)\gamma_n u(u-1) \dots (u-n+3) \Delta_x^{n-3}]$$

$$\Delta^n f(x) = \Delta_x^n [n! \gamma_n]$$

For abbreviation we shall write $\Delta^2 f(x) = \Delta_2^2$. Then

$$\Delta_2^{n-1} = \Delta_2^{n-2} - \Delta_2^1$$

and

$$f(\beta_1) = \gamma_0$$

$$\Delta_1 = \Delta_x \gamma_1$$

$$\Delta_2 = \Delta_x k! \gamma_2$$

$$\dots$$

$$\Delta_i^n = \Delta_x^n n! \gamma_n \text{ (for } i=1, 2, \dots) \text{ holds.}$$

So we get Newton's formula

$$f(x) = f(\beta_1) + \Delta_1^1 u + \frac{1}{2!} \Delta_2^2 u(u-1) + \dots + \frac{1}{n!} \Delta_n^n u(u-1) \dots (u-n+1)$$

$$= f(\beta_1) + \frac{\Delta_1^1}{\Delta_x} (x - \beta_1) + \frac{1}{2!} \frac{\Delta_2^2}{\Delta_x^2} (x - \beta_1)(x - \beta_2) + \dots + \frac{1}{n!} \frac{\Delta_n^n}{\Delta_x^n} (x - \beta_1) \dots$$

$$x - \beta_n,$$

The elements Δ_i^k can be calculated very easily by the following scheme —

$$\begin{array}{ccccccc}
 \lambda_1 & & & & & & \\
 & \Delta_1^1 & & & & & \\
 \lambda_2 & & \Delta_2^1 & & & & \\
 & \Delta_2^2 & & & & & \\
 \lambda_3 & & & \Delta_3^1 & & & \\
 & & & & \Delta_3^2 & & \\
 & & & & & \Delta_3^3 & \\
 & & & & & & \Delta_3^4 \\
 & & & & & & & \Delta_3^5 \\
 & & & & & & & & \Delta_3^6 \\
 \lambda_4 & & & & & & & & & \Delta_3^7
 \end{array}$$

The degrees of $f(x)$, $\Delta^1 f(x)$, $\Delta^2 f(x)$, ... are decreasing and the last one is a constant, so we can use the above scheme as a *for extrapolation* to get the value of $f(x)$ for every arbitrary integral value of x that means for every value $x = x_1 + k\Delta_1$ where k is an arbitrary integral number.

Example. Let $f(x)$ be of degree 4 and let $f(1) = 1$, $f(2) = 4$, $f(3) = 9$, $f(4) = 16$, $f(5) = 25$. In order to get $f(6)$ we use the scheme, calculating at first the numbers above the dotted line from the left to the right and then the numbers below the dotted line from the right to the left.

$$\begin{array}{ccccccc}
 1 & 3 & & & & & \\
 & & 1 & & & & \\
 3 & 4 & & 0 & & & \\
 & & 1 & & -5 & & 16 \\
 5 & 9 & & -5 & & 10 & \\
 & & -4 & & 5 & & 15 \\
 4 & 1 & & 1 & & 2 & \\
 & & 1 & & 30 & & \\
 5 & 2 & & 31 & & & \\
 & & & & & & \\
 6 & 6 & & & & &
 \end{array}$$

Hence $f(6) = 36$.

PART V

MATRICES. RESULTANTS.

§ 1. MATRICES.

In the first part of these lectures matrices have been used to solve (1.1) systems of linear equations and in § 14 and § 15 a few properties of matrices have been discussed. We will now consider the matrices in a more systematic manner.

Let K be an arbitrary field (see Part II § 2) let 0 be the null element 1 be the unit element of K , and let

$$a, b, \dots, f, c, f \text{ with and without indices of any kind} \quad (1.1)$$

be arbitrary elements of K . A = table of m rows and n columns.*

$$\begin{pmatrix} a_{11} & a_{12} & \dots & a_{1n} \\ a_{21} & a_{22} & \dots & a_{2n} \\ \vdots & \vdots & \ddots & \vdots \\ a_{m1} & a_{m2} & \dots & a_{mn} \end{pmatrix} \quad (a_{ij}) = A \quad (1.2)$$

has been called (see Part I § 6) a *matrix* and a_{ij} its *elements*. If $m=n$, this number will be said to be the *degree* of A , the general case can be reduced to the case $m=n$.

Let A, B, C, \dots be matrices of degree n and let the elements be denoted by the corresponding small latin type as in (1.2). The *addition* of matrices is given by

$$A + B = P, \text{ where} \quad (1.3)$$

$$a_{ij} + b_{ij} = p_{ij} \quad \text{for } i=1, \dots, n \quad j=1, \dots, n$$

$$\text{The commutative law} \quad A + B = B + A \quad (1.4)$$

$$\text{and the associative law} \quad (A + B) + C = A + (B + C) \quad (1.5)$$

hold for the addition of matrices.

Let c be an arbitrary element of the field K , then we define the *product*

$$cA = (ca_{ij}) \quad (1.6)$$

i.e., we multiply every element of A with c and get a new matrix cA of degree n . Then the

$$\text{distributive laws} \quad (c+d)A = cA + dA \quad (1.7)$$

$$c(A+B) = cA + cB$$

* Instead of brackets sometimes vertical double bars are put

hold. For $e=0$, we get

$$0A=0, \quad (1.8)$$

the entries of 0 being equal to 0 and

$$A+0=A \quad (1.9)$$

If we define the subtraction by

$$A-B=A+(-1)B, \quad (1.10)$$

then $(A+B)+C=A+(B+C)$ (see Part II [1.2]), M in which the elements are not added. In M the multiplication with an element of K is a linear transformation (see Part I [1.1]).

If $e=1$, M is a commutative K -algebra (see Part II [1.1]). M will be proved to be commutative by using the following definition. Let $E(p,k)$ be defined by

$$\begin{aligned} e_{ij}^p(r,k) &= 1 \\ e_{ij}^p(r,s) &= 0 \quad \text{for } (p,k) \neq (r,s) \end{aligned} \quad (1.11)$$

then
$$\sum_{i,j} e_{ij}^p E(p,k) = A \quad (1.12)$$

holds and A is invertible and only if every e_{ij}^p is equal to 0 . Hence the n^2 matrices are independent. They form a basis of M , and M can be considered as a vector space of rank n^2 , (see Part I [1.4] Part II [1.1]). By the theorem of Artin and Schreier the module is uniquely defined up to an arbitrary isomorphism.

[1.2] In Part I [1.1] the multiplication of matrices of M has been defined by

$$A \cdot B = D, \quad d_{ij}^p = \sum_k a_{ik}^p b_{kj}^p \quad (1.13)$$

and has satisfied law $(AB)C = A(BC)$ (1.14)

has also been proved. From (1.13) and (1.14) the

distributive laws
$$A+(B+C) = (A+B)+C$$

$$C(A+B) = CA+CB \quad (1.15)$$

follow directly.

Moreover for $n=2$ see Part II [1.2]. The reader may prove as an exercise that the ring is non commutative except when $n=1$. The ring has as the unit element, the matrix

$$E = e_{ij}^1 \quad \text{where } e_{ii}^1 = 1 \text{ and } e_{ij}^1 = 0 \text{ for } i \neq j \quad (1.16)$$

Then $E A = A E = A$. (1, 17)

and $O A = A O = O$ (1, 18)

for every matrix A . But it is possible that $A \neq I$ may be equal to A if $B \neq I$ or it may be equal to O if $B = I \neq A$.

$$\text{E.g. let } A = \begin{pmatrix} a & \\ & 0 \end{pmatrix} \text{ and } B = \begin{pmatrix} 1 & \\ & 1 \end{pmatrix} \text{ or } \begin{pmatrix} 1 & \\ & 0 \end{pmatrix}$$

To every matrix A there corresponds a determinant $\det A$ and (see Part I § 16)

$$\det (AB) = \det A \cdot \det B \quad (1, 19)$$

On the other hand, there exists a matrix A^{-1} and a matrix A for which $\det A = a \neq 0$ and $\det A^{-1} = 1/a$ and the other elements are equal to 0. This is the inverse of A in the group of matrices. In the convention $A^{-1} = \det A$ the set of all matrices is a group. In the notation of Part II, $A^{-1} = 1/A$ and it is clear that the multiplication remains invariant.*

$$A^{-1} A = I \quad \begin{pmatrix} 1 & & \\ & 1 & \\ & & 1 \end{pmatrix} \quad A A^{-1} = I$$

It is therefore possible to multiply a matrix A by an element c of K by the multiplication cA or Ac and it is mentioned that $\det(cA) = c^n \det A$.

$$\det(cA) = c^n \det A$$

* It is important to use with every matrix A

$$cB, C = bc, BC, \quad (1, 21)$$

If $\det A = 0$ and $\det A = a \neq 0$ then $AX = 0$ and $AX = I$ have the same solution $X = 0$ and $X = I$ respectively and let

$$L_i = ab^i \quad (1, 22)$$

be the cofactor of a_i (see Part I p. 21),

* The notation A^{-1} is used for the inverse of A in the group of matrices.

then $\sum a_j b_j = 0 = \sum a_i b_i$ for $i \neq k$

$\sum_{j=1}^n a_j b_j = 1 = \sum_{i=1}^n a_i b_i$ [cf. the sec. Part I (1) and (1')] Hence

$$AB = E = BA \quad (1, 22)$$

hence The matrix B (the elements of which have been defined by (1, 22)) is said to be the inverse of A and will be denoted by A^{-1} .

$$\text{So } AA^{-1} = E = A^{-1}A \quad (1, 23)$$

There exist therefore an inverse matrix if and only if the determinant is different from 0.

Let $\det A \neq 0$, then from

$$AX = B, \quad YA = B \quad \text{it follows that}$$

$$X = A^{-1}B, \quad Y = BA^{-1}$$

Hence the uniqueness with non-vanishing determinant follows from a set in which the two inverse operations of the multiplication are correct and give a unique result. The existence follows as the determinant of the n in the n matrices with non-vanishing determinant may be equal to 0. Of course every matrix is the sum of two matrices, each with a vanishing determinant, therefore may prove the proposition as in exercise.

[1/3] We have to consider now invariant matrices. To every matrix A there exists a linear transformation of the space of the n vectors over K transforming the n vector x_1, \dots, x_n into x'_1, \dots, x'_n :

$$a_1(x_1 + a_2x_2 + \dots + a_nx_n) = x'_1 \quad (1, 24)$$

$$i = 1, \dots, n$$

By this transformation the covectors over Part I p. 2 x^i have been transformed to the vectors $x'_i = [x^i] = [x^i] = [x^i]$ defined by the lineation. If we want to write formula (1, 24) as a matrix formula it is convenient to introduce a special notation for matrices in which every element outside the first column is equal to 0.

$$\text{Let } \begin{pmatrix} x_1 & 0 & \dots & 0 \\ x_2 & 0 & & 0 \\ \vdots & \vdots & \ddots & \vdots \\ x_n & 0 & & 0 \end{pmatrix} = e_i \quad (1, 25)$$

then (1.24) becomes

$$A(x) = (x^i) \quad (1.24)$$

$$(x) = A^{-1}$$

The formulae (1.24) and (1.25) express the transformation by which the basis of the n -vector-space formed by the unit-vectors e^i becomes transformed to the vectors $\sum a^i_j e^j$.

$$\text{Let } x = By, \quad y = B^{-1}x \quad \text{and } \det B \neq 0, \quad (1.26)$$

then

$$ABy = B(y')$$

$$B^{-1}AB y = (y'). \quad (1.27)$$

By (1.26) a linear (1-1) correspondence exists between the vectors of the vector-space which are set in (1.26) the unit-vectors of the system there correspond the vectors

$$(B_j) = (b^i_j), \quad (b^i_j)$$

of the system. These vectors form a basis and conversely each basis of the vector-space can be chosen as a set of vectors b^i_j as b^i_j defines the columns of a matrix B with non-vanishing determinant. The vectors $(b^i_j) = B_j$ become transformed by (1.27) in the same manner as the unit-vectors become transformed by (1.24) $\rightarrow \sum a^i_j e^j$. So (1.27) give the transformation of the vector-space if the transformation of an arbitrary basis is given.

The matrix

$$B^{-1}AB \quad (1.28)$$

is said* to be the *transform* of A by B A_B .

$$E^{-1}AE = A, \quad B = B^{-1}AB, \quad B^{-1}B = A, \quad C^{-1}B = AC, \quad C = (C^{-1})^{-1}AB, \quad$$

hold the transform of a fixed matrix A from a basis (see Fact II. 1.3) and from (1.19) it follows that the matrices of the same type have the same determinant. The transformations generated by different matrices of the same type are isomorphisms, they correspond to different bases of the vector-space. If we replace the unit-vectors by the basis formed by the vectors B_j (1.26) if we give the vector defined by x the coordinates x_1, \dots, x_n of $x = B_j x_j$, the matrix $C = B^{-1}AB$ becomes reduced by A . If it is given, we may arrange that A will be reduced to a normal form by a suitable choice of B .

* Some authors denote $B^{-1}AB$ as the transform of A by B^{-1} .



and if $\det A \neq 0$, $A^{-1} = (A^*)^{-1}$, then it follows that

$$A^* A^* = A^{-1*} = A^* A^*, \quad (2.6)$$

The powers of A are commutative matrices.

Let $\phi(x) = \sum_{j=0}^n a_j x^j$, a_j in \mathcal{A} , $a_n \neq 0$, $y(x) = u(x)A^{-1}$, $u(x)$ defined by $\phi(A)$ the matrix

$$\phi(A) = \sum_{j=0}^n a_j A^j = a_0 I + a_1 A + \dots + a_n A^n \quad (2.7)$$

and $\phi \neq 0$ and $\phi(A)$ is invertible in \mathcal{A} . Let $\psi(x)$ be a polynomial in $\mathcal{A}[x]$, $\psi(x) = \sum_{j=0}^m b_j x^j$, $b_m \neq 0$, and (2.8) it follows that

$$\phi(A)\psi(A) = \omega(A) = \psi(A)\phi(A), \quad (2.8)$$

Hence the integral functions $f(x) = \int_0^x \phi(A) \psi(A) dx$ are in \mathcal{A} . We cannot expect that $\phi(A)$ and $\psi(A)$ are matrices of degree n over \mathcal{A} as this ring is not necessarily a field. In fact, \mathcal{A} is a ring in $\mathbb{C}[x]$. In this ring there are exact n^2 independent matrices hence the commutative subring cannot contain more than n^2 independent matrices. The matrices

$$I, A, A^2, \dots, A^{n^2}$$

therefore cannot be independent, and there must be a polynomial $\omega(x)$ of degree $\leq n^2$, with the property that

$$\omega(A) = 0,$$

so the matrices may be considered as roots of a polynomial $\omega(x)$ in $\mathcal{A}[x]$ and let be the characteristic polynomial of A^{-1} $\chi_A(x)$ on \mathcal{A}_1^n of $\chi_A(x)$ there corresponds at least one λ but n $\neq (0, \dots, 0)$ of (2.1).

Theorem. If $\chi_A(x) = \chi_{A^{-1}}(x)$ are the solutions of $\chi_A(x) = 0$, $\chi_{A^{-1}}(x) = 0$ a solution of (2.9) corresponds to $\lambda = 0$ then the vectors $(\beta_1), \dots, (\beta_n)$ are independent.

Proof. Let the theorem not be true, then there are $\leq n$ vectors say β_1, \dots, β_n which are dependent but every minor must

of them is independent. Then an equation

$$a_1(\beta_1) + \dots + a_n(\beta_n) = 0 \quad (2.9)$$

holds $\forall \beta_1, \beta_2, \dots, \beta_n \in F$ if and only if $a_1 = a_2 = \dots = a_n = 0$. If A we get

$$a_1\lambda_1(\beta_1) + \dots + a_n\lambda_n(\beta_n) = 0.$$

Hence $a_1\lambda_1 + a_2\lambda_2 + \dots + a_n\lambda_n = 0$ as $\beta_1, \beta_2, \dots, \beta_n$ are independent elements.

Also $\lambda_1, \lambda_2, \dots, \lambda_n$ are linearly independent elements. Hence $a_1 = a_2 = \dots = a_n = 0$. Hence the β_i 's are linearly independent. Hence A is diagonalizable.

Let A be a matrix over F . Then A is diagonalizable if and only if A can be

transformed into the diagonal-matrix

$$\begin{pmatrix} \lambda_1 & 0 & \dots & 0 \\ 0 & \dots & \dots & \dots \\ 0 & \dots & 0 & \lambda_n \end{pmatrix}. \quad (2.10)$$

Let $\beta_1, \beta_2, \dots, \beta_n \in F$ be linearly independent. Hence the matrix $B = (\beta_1, \beta_2, \dots, \beta_n) \in F^{n \times n}$ is invertible. Hence we can transform A into a diagonal matrix D by $B^{-1}AB = D$. Hence A is diagonalizable.

If n is not a prime, then the above result does not hold in every

case. **Ex. 2.** Let $A = \begin{pmatrix} 1 & 1 \\ 0 & 1 \end{pmatrix}$. Then $\lambda_A(x) = 1 - x^2$, $\lambda_1 = \lambda_2 = 1$.

$A - I = \begin{pmatrix} 0 & 1 \\ 0 & 0 \end{pmatrix}$ has the rank 1, hence $A - I$ has only one solution

$(x_1, x_2) = (1, 0)$.

Let $R = k[x]$ be the ring of polynomials over k , let ϕ be a homomorphism from R to the ring $M_n(k)$. Then $\phi(x) = A$ is a matrix in $M_n(k)$, although A is not invertible. From $\phi(x^2) = A^2 = 0$, $\phi(x) \neq 0$, it follows that $\phi(x) = 0$. Hence the number of the matrices which are roots of a polynomial may be greater than the degree of the polynomial. Even it may be infinite.

Ex. 3. Let $A, B \in M_n(k)$. Then $A^t B^t = (B A)^t = B^t A^t$ holds.

Ex. 4. Let $A, B \in M_n(k)$. Then $(B^{-1} A B)^t = B^{-1} A^t B$ holds.

Hence from $\phi(A) = 0$ it follows that $\phi(B^{-1} A B) = 0$ where B is an arbitrary matrix with non-vanishing determinant. In other words



Theorem 1. Two similar matrices have the same eigenvalues.

A similar theorem holds for the characteristic polynomial.

Theorem 2. Two similar matrices have the same characteristic polynomial. [2.11]

Proof. Let $B^{-1}AB = C$, $B^{-1} = B^{-1}$, $B = B^{-1}B$, $B^{-1}B = I$.

$$\det(B^{-1}AB) = \det A \quad \text{holds}$$

$$B^{-1}(A - xE)B = B^{-1}AB - xE = C - xE; \quad \text{hence}$$

$$\det(A - xE) = \det B^{-1}(A - xE)B = \det(B^{-1}AB - xE) = \det(C - xE).$$

$$\text{i.e., } \chi_A(x) = \chi_{B^{-1}AB}(x).$$

The two last theorems have a certain extension to them. It will be proved later on that every matrix is similar to a block diagonal matrix. Now we will consider only the case when the eigenvalues of the matrix and its characteristic polynomial are all different.

$$A - \lambda_1 E = 0, \quad (A - \lambda_1 E)(\beta_1) = 0. \quad \text{As the } n \text{ matrices}$$

$$A - \lambda_1 E, \beta_1 \text{ are equivalent, } \lambda_1 = \lambda_1, \quad \text{if } A - \lambda_1 E = 0, \quad \beta_1 = 0.$$

The vectors $\beta_1, \beta_2, \dots, \beta_k$ form a basis of the subspace of rank k of the vector space of rank n is transformed by $A - \lambda_1 E$ into a subspace of rank k .

Hence (see Part I § 14):

$$\chi_A(A) = 0. \quad (2.11)$$

Let χ_A be a multiple root of the characteristic polynomial of A . Then it is a multiple root of the same characteristic polynomial of $B^{-1}AB$ where B is an arbitrary matrix of $(1 \times n)$ and $n \times 1$ vanishing determinant. If A there corresponds to the eigenvalue λ_1 of (2.11). Let β_1 be the first column of a matrix B then the matrix $B - \lambda_1 E$ is transformed by $B^{-1}AB$ into the same manner as $B - \lambda_1 E$ hence $B^{-1}AB - \lambda_1 E = 0$. If $B - \lambda_1 E = 0$.

Therefore

$$B - \lambda_1 E = 0 \quad \left| \begin{array}{c} \lambda_1 \\ \vdots \\ \lambda_1 \end{array} \right| \quad (2.12)$$

In this form with asterisks * denote certain elements which will not be considered particularly and λ_1 is a number of given $n - 1$.

$\lambda = \chi_A(x) = x^r \chi_{A_1}(x) \dots \chi_{A_{r-1}}(x) = \lambda_1(x) \chi_{A_1}(x) \dots \chi_{A_{r-1}}(x)$ is a root of $\chi_{A_1}(x)$ of order $r-1$. We therefore can transform A' into

$$B'A'B'^{-1} = \begin{pmatrix} \lambda_1 & \dots & 0 \\ 0 & \dots & 0 \\ \vdots & & \vdots \\ 0 & \dots & 0 \end{pmatrix} \begin{pmatrix} A'' & & \\ & \ddots & \\ & & A'' \end{pmatrix}$$

where A'' is of degree $n-r$, and if

$$B_1 = \begin{pmatrix} 1 & & 0 & \dots & 0 \\ & 1 & & & \\ & & \ddots & & \\ & & & 1 & \\ 0 & & & & 1 \end{pmatrix} = B_2, \quad (2.13)$$

$$B_2 A B_2^{-1} = \begin{pmatrix} \lambda_1 & \dots & 0 & \dots & 0 \\ & \lambda_1 & \dots & & \\ & 0 & \dots & & \\ & 0 & \dots & & \\ & \vdots & & & \vdots \\ & 0 & \dots & & 0 \end{pmatrix} \begin{pmatrix} A'' & & \\ & \ddots & \\ & & A'' \end{pmatrix}$$

The first row of $B_2 A B_2^{-1}$ is $\lambda_1 \dots \lambda_{r-1} \lambda_1 \dots \lambda_{r-1} \lambda_1$ (the second row is $\lambda_1 \dots \lambda_{r-1} \lambda_1 \dots \lambda_{r-1} \lambda_1$), and so on. If $r=2$, then λ_1 is a root of $\chi_A(x) = (x-\lambda_1)^n$, and so we continue the procedure till we get

$$B_1^{-1} A B_1 = \begin{pmatrix} \lambda_1 & \dots & 0 & \dots & 0 \\ \lambda_1 & \dots & 0 & \dots & 0 \\ \vdots & & \vdots & & \vdots \\ 0 & \dots & \lambda_1 & \dots & 0 \\ 0 & \dots & 0 & \dots & \lambda_1 \\ \vdots & & \vdots & & \vdots \\ 0 & \dots & 0 & \dots & 0 \end{pmatrix} \begin{pmatrix} A'' & & \\ & \ddots & \\ & & A'' \end{pmatrix} \quad (2.14)$$

A'' is a matrix of degree $n-r$, $\chi_{A''}(x) = \chi_A(x) - (x-\lambda_1)^r$ is not divisible by $(x-\lambda_1)^r$. The r matrices B_1, B_2, \dots, B_r are transformed by A in the same manner as the matrices B_1, B_2, \dots, B_r are transformed by $B_1 A B_1^{-1}$. Hence

$$\begin{aligned} A(\beta_1) &= \lambda_1 \beta_1 \\ A(\beta_2) &= \lambda_1 \beta_2 + c(\beta_1) \\ A(\beta_3) &= \lambda_1 \beta_3 + d_1(\beta_1) + d_2(\beta_2) \\ &\vdots \\ A(\beta_r) &= \lambda_1 \beta_r + h_1(\beta_1) + \dots + h_{r-1}(\beta_{r-1}). \end{aligned}$$

Therefore

$$(A - \lambda E)(\beta_1) = 0 \quad (2, 15)$$

$$(A - \lambda E)(\beta_2) = c(\beta_1), \quad (A - \lambda E)^2(\beta_2) = 0$$

$$(A - \lambda E)(\beta_3) = d_1(\beta_1) + d_2(\beta_2), \quad (A - \lambda E)^2(\beta_3) = d_2c, \quad (A - \lambda E)^3(\beta_3) = 0$$

$$(A - \lambda E)^3(\beta_3) = 0,$$

Hence for every vector α of the vector space V generated by $\alpha_1, \dots, \alpha_{r-1}, \beta_1, \dots, \beta_r$

$$(A - \lambda E)^r(\alpha) = 0 \quad (2, 16)$$

holds. The vectors α_i are columns of a matrix with non-vanishing determinant, hence they are independent. Therefore $\text{rank } A = r$.

We want to prove now that every vector γ , for which

$$(A - \lambda E)^r(\gamma) = 0 \quad (2, 17)$$

holds, belongs to V . If there would be such a vector γ outside of V , we can choose it so that $(A - \lambda E)^{r-1}\gamma$ is a vector of V . If

$$A(\gamma) = \lambda(\gamma) + h_1(\beta_1) + \dots + h_r(\beta_r),$$

As β_1, \dots, β_r are supposed to be independent, there exists a matrix C , of which these vectors form the $r+1$ last rows

Then

$$(A - \lambda E)^{r+1} \begin{pmatrix} \lambda & 0 & \dots & 0 \\ 0 & \lambda & \dots & 0 \\ \vdots & \vdots & \ddots & \vdots \\ 0 & 0 & \dots & \lambda \\ 0 & 0 & \dots & 0 & A^{r+1} \\ 0 & 0 & \dots & 0 \end{pmatrix}$$

and therefore $\chi_A(x) = (A - x)^{r+1} \chi_A(x)$, x contrary to the supposition that r is the highest exponent of $A - x$ in $\chi_A(x)$. The vector space (2, 17)

is therefore composed of all vectors which satisfy (2, 16) for any exponent r . As $(A - \lambda E)(\alpha) = (\alpha')$ satisfies (2, 17) if α satisfies it, $A(\alpha) = (\alpha') + \lambda(\alpha)$ is a vector of V if α belongs to V . By these considerations we get the following theorem.

Theorem. If λ is a root of $\chi_A(x)$ of order r , then the vectors satisfying (2, 17) for any q form a vector space V of rank r . Each vector α of V satisfies (2, 16) and is transformed by A to a vector of V . If the first r columns of B form a suitably chosen basis of V , then (2, 14) holds

[2/5] Let

$$\chi_A(x) = (\lambda_1 - x)^{r_1} \dots (\lambda_m - x)^{r_m}, \quad (2, 18)$$

where $\lambda_1, \dots, \lambda_m$ are distinct. To every λ_i there corresponds a vector space V_i of rank r_i so that for every vector (a_i) of V_i , the equation $(A - \lambda_i E)^{r_i}(a_i) = (0)$ holds and (a_i) is transformed by A into a vector of V_i . To prove that the vectorspaces V_i are independent, we have to use the following lemma.

Lemma. The k -th coefficient of the polynomials $\phi_1(x), \dots, \phi_m(x)$ of $V_1(x)$ can be expressed by $\phi(x) = \phi_1(x)\psi_1(x) + \dots + \phi_m(x)\psi_m(x)$ where ψ_1, \dots, ψ_m are polynomials of $\Lambda[x]$.

Proof. The lemma is true if $m = 1$ and $m = 2$ (see Part II 4/5); let it be true for $m \leq k-1$ and therefore the k -th coefficient of $\phi_1, \dots, \phi_{k-1}$ be $\omega_1 = \phi_1(x)\omega_1(x) + \dots + \phi_{k-1}(x)\omega_{k-1}(x)$. As the k -th coefficient of ϕ_1, \dots, ϕ_k is the k -th coefficient of ϕ and ϕ , there is $\phi(x) = \omega_1(x)\psi_1(x) + \dots + \phi_m(x)\psi_m(x) = \sum \phi_i(x)\psi_i(x)$.

The preceding lemma will be applied to the polynomials

$$\phi_i(x) = \chi_A(x) : (x - \lambda_i)^{r_i},$$

As these polynomials are relatively prime, there exist m polynomials $\eta_i(x)$ satisfying

$$\eta_1(x) + \dots + \eta_m(x) = 1 \quad (2, 19)$$

$\eta_i(x)$ is divisible by $(x - \lambda_k)^{r_k} = 1, \dots, m$ and therefore

$$\text{by } (x - \lambda_k)^{r_k} \quad \text{for } k \neq i.$$

Hence $\eta_i(A)(a_i) = (0)$ and from (2, 19) it follows that $\eta_i(A)(a_i) = (a_i)$. (2, 20)

It follows that $\eta_i(A)(a_i) = (a_i)$.

If n vectors a_1, \dots, a_n of the different vectorspaces V_i satisfy

$$(a_1) + \dots + (a_m) = (0),$$

we get by multiplication with the matrices $\eta_i(A)$ for every i

$$\eta_i(A)(a_i) = (a_i) = (0).$$

Hence the vectorspaces V_1, \dots, V_m are independent.



If $\{v_1, \dots, (v_{r_1}^f)^f, \dots, (v_{r_m}^f)^f\}$ is a basis of V , the $n = r_1 + \dots + r_m$ vectors

$$\{v_1\}, \dots, (v_1^f)^f, \{v_2\}, \dots, \dots, (v_m^f)^f \quad (2.21)$$

form a basis of the space of the n vectors. The matrix C whose columns are the vectors (2.21) has therefore a determinant $\neq 0$. The vector-spaces V_i are invariant for the transformation A . Hence the n vector-spaces generated by

$$\{v_1\}, \dots, (v_1^f)^f, \{v_2\}, \dots, (v_2^f)^f, \dots,$$

$$\{v_{r_1+1}\}, \dots, (v_{r_1+1}^f)^f, \dots, (v_{r_1+r_2}^f)^f,$$

$$\{v_{n-r_m+1}\}, \dots, (v_{n-r_m+1}^f)^f, \dots, (v_n^f)^f$$

are invariant for $C^{-1}AC = J$. This matrix has therefore the form

$$C^{-1}AC = \begin{pmatrix} A_1 & & \\ & \ddots & \\ & & A_m \end{pmatrix} \quad (2.22)$$

where A_i is a matrix of degree r_i and its characteristic polynomial is equal to $(\lambda_i - \lambda)^{r_i}$. Every element situated outside the square is equal to 0. As it has been proved in [2.4], the equation $(\lambda - \lambda_i - 1)(\lambda - \lambda_i) = 0$ holds for every vector $(v_i)^f$ of V_i . As the vectors (2.21) form a basis of the total vector-space, every vector v is the sum of vectors v_i . Hence

$$(\lambda_i - \lambda)(\lambda - \lambda_i - 1)A_i(v_i)^f = 0 \quad \text{if } A_i(v_i)^f = v_i \quad \text{holds for every vector } v_i.$$

Hence $\chi_{A_i}(A) = 0$.

The cause of these considerations is given by the following theorem.

Theorem. Let $\chi(\lambda)$ be the characteristic polynomial of A , then there exist m independent vector-spaces V_i of rank r_i , $i = 1, \dots, m$. These vector-spaces are invariant for A . If the columns of C are basis

every vector space is included in the subsequent spaces. Let

$$u_{r-1}, \quad u_{r-2}, \quad \dots, \quad u_1, \quad u_0$$

be the rank of the vector spaces

$$W_{1, r-1}, \quad W_{1, r-2}, \quad \dots, \quad W_{1, 1}, \quad W_1$$

then there exists a basis

$$(\beta_1^1), \dots, (\beta_{u_{r-1}}^1), (\beta_1^2), \dots, (\beta_{u_{r-2}}^2), \dots, (\beta_1^r), (\beta_{u_1+1}^r), \dots, (\beta_{u_0}^r)$$

of W_1 with the property that

$$(\beta_1^i), \dots, (\beta_{u_{r-i}}^i)$$

is a basis of $W_{1, r-i}$ for every index i . (2.19)

Let $u_r \leq i < u_{r-1}$, then there exists a vector (β_1^{r+1}) satisfying

$$(A - \lambda E)^2 (\beta_1^{r+1}) = (\beta_1^r) \quad (2.20)$$

and we define (β_k^{r+1}) , for $1 \leq k \leq r+1$ by

$$(A - \lambda E)^{r+2-k} (\beta_k^{r+1}) = (\beta_k^r) \quad (2.20')$$

From (2.20') it follows that (2.20') holds for $k = 1$.

Let the vectors (β^i) be arranged in a triangular scheme

$$(\beta_1^1) \quad (\beta_{u_{r-1}}^1) \quad (\beta_{u_{r-1}+1}^1) \quad \dots \quad (\beta_{u_{r-2}}^1) \quad (\beta_{u_{r-2}+1}^1) \quad (\beta_{u_{r-1}+1}^2) \quad \dots \quad (\beta_{u_0}^2)$$

$$(\beta_1^2), \dots, (\beta_{u_{r-1}}^2), (\beta_{u_{r-1}+1}^2), \dots, (\beta_{u_{r-2}}^2), \dots, (\beta_{u_1+1}^r), \dots, (\beta_{u_0}^r) \quad (2.21)$$

$$(\beta_1^r), \dots, (\beta_{u_{r-1}}^r)$$

If we multiply a vector of this scheme from the left by $(A - \lambda E)$ matrix $(A - \lambda E)$ we get the vector just above it

$$(A - \lambda E) (\beta_1^{r+1}) = (\beta_1^r) \quad \text{hence}$$

$$\lambda (\beta_1^{r+1}) = \lambda (\beta_1^{r+1}) + (\beta_1^r) \quad \text{holds}$$

The vector space U , generated by the vectors of an arbitrary column (say the i -th), is therefore invariant under the transformation by A . These vectors

are independent, viz., from

$$(0) = c_1(\beta_1^t) + \dots + c_s(\beta_s^t) \quad \text{it follows that}$$

$$(0) = (\lambda - \lambda E^{s+1}, \lambda^t) \cdot \begin{pmatrix} \beta_1^t \\ \vdots \\ \beta_s^t \end{pmatrix} \quad \text{hence } c_s = 0.$$

The t -th row by itself forms a basis at U^t and in terms of this basis the transformation T^t is given by the matrix

$$\begin{pmatrix} \lambda & 1 & 0 & \dots & 0 \\ 0 & \lambda & 1 & \dots & 0 \\ \vdots & \vdots & \vdots & \ddots & \vdots \\ 0 & 0 & \dots & \lambda & 1 \\ 0 & 0 & \dots & 0 & \lambda \end{pmatrix} \quad (2, 28)$$

where the diagonal elements are λ , in the adjacent parallel one the elements are 1 and the other elements are 0. The degree of (2, 28) is equal to $s+1$ where s is given by (2, 27). To get the normal form of the transformation A we have to prove that the vectors (2, 27) are independent and form a basis of V . The vectors of the first r rows are independent and form a basis of W_1 by definition.

Let $\sum_{i=1}^{s+1} c_i \beta_i^t = 0$ where β_i^t is any vector of W_1 then we get by

multiplication with $(\lambda - \lambda E^{s+1})$ that $\sum_{i=1}^{s+1} c_i \beta_i^t = 0$ but as the vectors (β_i^t) are

independent, $0 = c_1 = \dots = c_{s+1}$ holds, and therefore $\beta_i^t = 0$.

Hence the vectors of the two first rows are independent. By mathematical induction it follows* that the vectors (2, 27) are independent.

$$(\lambda - \lambda E^{s+1}) \beta^t = \sum_{i=1}^{s+1} (\lambda - \lambda E^{s+1}) \beta_i^t, \text{ as it is a vector of } W_{1,1}$$

Let $(\beta^t = \sum_{i=1}^{s+1} \lambda_i \beta_i^t = \gamma)$ then $(\lambda - \lambda E^{s+1}) \gamma = 0$ hence γ is a

vector of $W_{1,1}$ i.e.

$$(\gamma) = \sum_{i=1}^{s+1} k_i (\beta_i^t)$$

* The above induction is the one-formal induction of an ascending

β^2 is therefore dependent on the vectors of the two first rows. Hence these vectors form a basis of W_2 . If β^2 is a vector of $W_{2,1}$, then $\beta^2 = A(AE - \beta^1) = A(AE - I)\beta^1 = \Sigma I\beta^1$ belongs to $W_{1,1}$, hence $k_2 \geq k_1 = 0$.

$(A - AE)\beta^2 = A(AE)^2\beta^1$ belongs to $W_{2,1}$, hence $k_2 \geq k_1 = 0$.

The vectors $(\beta^1), \dots, (\beta^{k_1}), (\beta^{k_1+1}), \dots, (\beta^{k_2})$ form therefore a basis of the vector space $W_{2,1}$.

By mathematical induction it follows that the vectors of the first k rows form a basis of W_k , and if we add the vectors $(\beta^{k+1}), \dots$ we get a basis of W_{k+1} . So we get that for $k = r$, the vectors (2.27) form a basis of the total vector space V .

We arrange this basis according to the columns, and we transform in such a manner that the vectors of this basis become unit vectors. Then A will be transformed into

$$\begin{array}{c}
 A^1 \\
 A^2 \\
 \vdots \\
 A^r
 \end{array}
 \quad
 \begin{array}{c}
 \vdots \\
 \vdots \\
 \vdots \\
 \vdots
 \end{array}
 \quad
 \begin{array}{c}
 C A C^{-1} = \overline{A} \\
 \vdots \\
 \vdots \\
 \vdots \\
 \vdots
 \end{array}
 \quad
 \begin{array}{c}
 \vdots \\
 \vdots \\
 \vdots \\
 \vdots
 \end{array}
 \quad
 \begin{array}{c}
 A^{*1} \\
 A^{*2} \\
 \vdots \\
 A^{*r}
 \end{array}
 \quad
 (2, 28)$$

The transformation of the subspaces U_i is given by A^i . Hence these matrices have the normal form (2.28). The degree of the matrix A^i is equal to the length of the corresponding column in (2.27). The degrees form a monotonically decreasing sequence. On the other hand if a matrix is given in the normal form (2.28) (2.29), the ranks of the subspaces W_1, W_2, \dots can easily be found out. The rank of W_k being the number of the matrices A^i , of which the degree is not less than k .

Different normal forms therefore belong to different classes of matrices. In the general case we can transform each matrix A , of (2.24) into its normal form. By these considerations the transforming theorem has been proved.

* See p. 58, footnote.

Let (2, 18) be the characteristic polynomial of an arbitrary matrix A . Then A can be transformed into a normal form (2, 22). The matrices A_1, \dots, A_n have the form (2, 23) with the degrees of the A_1', \dots, A_n' form a non-increasing sequence and each of the n has the form (2, 24) with $\alpha_i = 1$ for $i = 1, \dots, n$. The A_1, \dots, A_n remains arbitrary.

§1. SOME PROPERTIES OF THE NORMAL FORM AND OF THE CHARACTERISTIC POLYNOMIAL

Let A be a matrix with entries in a field F and let λ be an element of F . The matrix $A - \lambda E$ has the same determinant as A if $\lambda = 0$. If $\lambda \neq 0$, then $A - \lambda E$ is similar to A and hence has the same characteristic polynomial as A .

[9, 1] Let A be a matrix with entries in a field F and let λ be an element of F . The matrix $A - \lambda E$ has the same determinant as A if $\lambda = 0$. If $\lambda \neq 0$, then $A - \lambda E$ is similar to A and hence has the same characteristic polynomial as A .

The normal form of a matrix with entries in a field F is unique. The matrix (2, 22) is the normal form of A if and only if it is similar to A and if it is in normal form.

$$\begin{pmatrix} \gamma_1 & \gamma_2 & \dots & \gamma_r \\ & \gamma_1 & \dots & \gamma_{r-1} \\ & & \dots & \gamma_1 \\ & & & \gamma_1 \end{pmatrix} \quad (9, 1)$$

where $\gamma_1 \neq 1$ and the elements γ_i on the diagonal are equal to 0. Hence (9, 1) is commutative to A and there is no other matrix commutative to (9, 1) than the matrices (9, 1) as every other transformation will not give a matrix of the same type corresponding to the normal form.

If $\gamma_1 = 1$ each of the matrices A_i is of degree 1. Hence $0 = \gamma_1 - 1 = 0$. In this case the vector space is identical with W . There are n spaces W_1, \dots, W_n , every basis of the vector space leads to the normal form. The normal form is therefore commutative to every matrix with determinant $\neq 0$. Of course it is a diagonal matrix with the diagonal elements $= \lambda$. In the general case the investigation of the admissible transformations of the case $\gamma_1 = 1$ the investigation of the matrices commutative to the normal form is more complicated and we

will not go into the details here. The more general problem of finding out the matrices commutative to an arbitrary real matrix A , can easily be reduced to that problem if A and B are commutative. If $A \in \mathbb{R}^{n \times n}$ and $C \in \mathbb{R}^{n \times n}$ are commutative too.

The matrix A is a root of the characteristic polynomial $\chi_A(x)$, but it may be that A is a root of a polynomial of lower positive degree.

Let

$$\chi_A(x) = (A - \lambda_1 E)^{r_1} (A - \lambda_2 E)^{r_2} \dots (A - \lambda_m E)^{r_m}$$

To $A - \lambda_1 E^{r_1}$ there corresponds the minor matrix A_1 in the normal form (2.22), and this minor matrix has the form (2.22) composed by diagonal matrices $A_1^1, \dots, A_1^{r_1}$. Let r_1 be the degree of A_1^1 , then $1 \leq r_1 \leq r_1$ holds and every vector α_1 of the vector space corresponding to A_1 satisfies the condition

$$(A - \lambda_1 E)^{r_1} \alpha_1 = (0)$$

Corresponding to this definition we obtain the set $\alpha_1, \dots, \alpha_{r_1}$ then

$$1 \leq r_1 \leq r_1 \quad \text{and}$$

$$(A - \lambda_1 E)^{r_1} \alpha_{r_1} = (0) \quad \text{hold,}$$

for every vector α_i of the vector space corresponding to A_1 .

$$\text{Let } \psi(x) = (x - \lambda_1)^{r_1} \dots (x - \lambda_m)^{r_m} \quad (3.2)$$

$$\text{then } \psi(A) \alpha_i = (0)$$

holds for every vector α_i of the vector space and therefore holds for every vector of the vector space.

$$\text{Hence } \psi(A) = 0 \quad (3.3)$$

$\psi(x)$ is a factor of $\chi_A(x)$ and is of lower positive degree except in the case when $r_1 = r_1$ for every k which means that the normal form (2.22) of every k shows only one matrix, $\alpha_1, \dots, \alpha_{r_1}$ and $r_1 = r_1$.

To prove that A is not a root of a polynomial which is not divisible by $\psi(x)$, we consider the case $\chi_A(x) = (A - \lambda_1 E)^{r_1}$. Let $\phi(x) = x - \lambda_1$ and let $\phi(x)$ not divide $\psi(x)$ and $\phi(x)$ is smaller than the degree of $\psi(x)$. Then $\phi(A) = A - \lambda_1 E \neq 0$ which is not true. Hence the transformation of the vector space generated by $\alpha_1, \dots, \alpha_{r_1}$ is an automorphism.

and there exists therefore a vector (x, y) that satisfies $(B/y) = x > 0$. Hence $\phi(A - I) \neq 0$. Therefore $\phi(A) \neq 0$.

Let $A = \begin{pmatrix} B & 0 \\ 0 & C \end{pmatrix}$. From the law of multiplication of matrices it

follows directly that $A^2 = \begin{pmatrix} B^2 & 0 \\ 0 & C^2 \end{pmatrix}$, $A^3 = \begin{pmatrix} B^3 & 0 \\ 0 & C^3 \end{pmatrix}$, and

therefore $\omega(A) = \begin{vmatrix} \omega(B) & 0 \\ 0 & \omega(C) \end{vmatrix}$ for every polynomial ω .

Hence $\omega(A) = 0$ if and only if $\omega(B) = \omega(C) = 0$.

The corresponding rule holds if A is composed of more than two diagonal matrices. If therefore A has the form $(2, 2, \dots, m)$, $\omega(A) = 0$ holds if and only if $\omega(A_i) = 0$ for $i = 1, \dots, m$.

Hence we must be able to say every $x = \lambda^{-1}$. By these considerations the following theorem has been proved:

Theorem. $\omega(A) = 0$ if and only if (x) is divisible by the polynomial $\chi_A(x)$, which has been defined by (3.2).

[3/3] By the *normal form* of A the transformation generated by A becomes uniquely defined only if the roots of the polynomial are all different. If there are equal roots, there are different normal forms and therefore different non-similar transformations corresponding to the same characteristic polynomial.

To get the characteristic polynomial, it is not necessary to put the matrix in the normal form. As

$$\chi_A(x) = \det(A - xB),$$

the coefficient of x^k in $\chi_A(x)$ is

$$(-1)^k \sum A_{n-k+1,1}, \quad (3.4)$$

where $A_{n-k+1,1}$ denote the minors of A with $n-k+1$ rows which are symmetric to the diagonal of A . The signs are according to the transformation;

the determinant of any one of these matrices is equal to $k \neq 0$ and $k = n - 1$.

$$\det(A) \neq 0 \quad \text{and} \quad \Sigma a_j \neq 0. \quad (3.4')$$

We want to apply the theory to the linear substitution of a complex variable

$$z = \frac{\alpha + \beta}{\gamma + \delta} \quad \text{where } \alpha, \beta, \gamma, \delta \neq 0.$$

We introduce homogeneous coordinates $z = z_1/z_2$, $\bar{z} = \bar{z}_1/\bar{z}_2$,

$$w_1 = \alpha z_1 + \beta z_2$$

$$w_2 = \gamma z_1 + \delta z_2$$

As a common factor of $\alpha, \beta, \gamma, \delta$ is arbitrary we can arrange that $\alpha\beta\gamma\delta \neq 0$ and may add a common factor $\neq 0$ arbitrary.

$$\text{Let } \begin{pmatrix} \alpha & \beta \\ \gamma & \delta \end{pmatrix} = A = \lambda_1 e^{i\varphi_1} e^{i\varphi_2} e^{i\varphi_3} + 1 = \lambda_1 e^{i\varphi_1} (\lambda_2 e^{i\varphi_2} e^{i\varphi_3})$$

Hence $\lambda_1 \lambda_2 = 1$, $\lambda_1 = re^{i\varphi}$, $\lambda_2 = r^{-1}e^{-i\varphi}$.

$$(1) \lambda_1 \neq \lambda_2, \text{ normal form } \begin{pmatrix} e^{i\varphi} & 0 \\ 0 & e^{-i\varphi} \end{pmatrix} \quad (3.5)$$

As a factor $\neq 0$ and a permutation of λ_1, λ_2 remain arbitrary, we can choose $1 \leq r$, $0 \leq \varphi < \pi$.

2) $\lambda_1 = \lambda_2 = \pm 1$, $\varphi = \pm 2\pi$, by a suitable choice of the common factor $\neq 0$ we can arrange that $\alpha = \delta$, hence $\lambda_1 = \lambda_2 = 1$.

There are two normal forms

$$\begin{pmatrix} 1 & 0 \\ 0 & 1 \end{pmatrix} \quad \text{and} \quad \begin{pmatrix} 1 & 1 \\ 0 & 1 \end{pmatrix} \quad (3.6)$$

These transformations are largely discussed in the elements of the theory of functions. The classes of transformations with the normal form (3.6) are said to be *linear* and especially for $r=1$ they are called *elliptic*, for $r \neq 1$ $\varphi \neq 0$ they are called *hyperbolic*. The first matrix (3.6) denotes the identity, the second matrix denotes a *parabolic displacement*, the infinite point of the complex sphere being the only fixpoint and the other transformations of this class are called *parabolic transformations*, the only fixpoint being a finite point.

§ 4. THEORY OF ELEMENTARY DIVISORS

[4.1]

Let S be an integral domain with the following properties:

1. $a \neq 0$ for every a in S ; if $a \neq 1$ then a corresponds to an integral number $\nu(a) > 1$.

2. If b is a factor of a , then

$$\sum_{i=1}^n \nu(b_i) \leq \nu(a) \quad (4.2)$$

where the $\nu(b_i)$ are functions of b_i and $\nu(a)$ is associated with a .

If a_1 is an arbitrary element of S and a_1 is not divisible by a_2 then there exist in S elements b and c , satisfying

$$b + ca = a_1, \quad N(a_1) < N(a_2) \quad (4.3)$$

It must be borne in mind that we have to be considered in Part II, § 4, and [4.1] of these conditions. In addition, these $\nu(a)$ are

1. The ring of integral numbers if we define N for a by $N(a) = |a|^2$ (definition of the ring of integers) and $\nu(a) = 1$.

2. The ring of the integral complex numbers $a + bi$ (see Part II, § 11), where a and b are integers. We define N for $a + bi$ by $N(a + bi) = a^2 + b^2$. The units of this ring are $\pm 1, \pm i, \pm 1 - i, \pm 1 + i$.

3. The ring $K[x]$ of the polynomials in an arbitrary indeterminate x over an arbitrary field K . We define N for every polynomial $f(x)$ which is different from the polynomial 0 by $N(f) = \text{degree}(f) + 1$. The units of this ring are the elements $\neq 0$ of K .

It has been proved in Part II, § 4. The factorisation in S is unique, and the h.c.f. of two elements a and b can be expressed by

$$(a, b) = c a + d b \quad (4.4)$$

where c , and d are elements of S .

As we see from the definition of [4.5] the h.c.f. of elements of S can be expressed by

$$(a_1, \dots, a_n) = c_1 a_1 + c_2 a_2 + \dots + c_n a_n \quad (4.5)$$

Exercise. Given an integral domain S for which (4.1), (4.2) and (4.3) hold, prove that it is possible to replace the function N by a function \bar{N} satisfying the same conditions as N , so that $\bar{N}(u) = 1$ if and only if u is a unity of S .

We proceed as the method of "sweep out" (see Part I § 6) [4/2] matrices with elements from Δ . The method is easily modified for Δ owing to the fact that in Part I the elements of the matrices were assumed to be real numbers. For complex elements of Δ the operation of "dividing" by an element of Δ always is possible. In this case we have to deal with the elements taken from a row, the division has therefore to be replaced by the adjunction of the factor. The subsequent steps of the "sweep out" of the row differ of course. They can be considered as operations with rows and columns and as dealing with the left side or from the right side of the matrix. It is not difficult to consider both interpretations at any step. The matrices we have to consider are the same as in Part I § 16.

$$\text{Diagonal matrix } D = (d_{ij}) \quad \text{with } d_{ii} = \lambda_i, \text{ and} \\ \text{for } i \neq k, d_{ik} = 0; \quad (4.3)$$

$$\text{Elementary matrix } E = (e_{ij}) = \begin{cases} e_{ii} = 1 & \text{here } i = k, \\ e_{ik} = \lambda, & \\ e_{ki} = \lambda, & \\ \text{every other } e_{ij} = 0 \end{cases} \quad (4.4)$$

We get

$$\begin{aligned} DA & \text{ by multiplying every row of } A \text{ with } d_{ii} = \lambda_i, \text{ and } d_{ii} = \lambda_i \\ AD & \text{ as } d_{ii} = \lambda_i \text{ column } A \text{ is } \lambda_i \text{ column } A \text{ is } \lambda_i \text{ column } A \text{ is } \lambda_i \\ E_{ik}(\lambda)A & \text{ by row } i \text{ of } A \text{ replacing the row } i \text{ by } \lambda y_i + y_k \\ AE_{ik}(\lambda) & \text{ by column } k \text{ of } A \text{ replacing the column } k \text{ by } \lambda y_k + y_i \end{aligned}$$

We consider especially the diagonal matrices

$$U = (u_{ij}^{\pm}) \quad (4.5)$$

for which the diagonal elements $u_{ii}^{\pm} = u_i$ are unities. By replacing each u_i by u_i^{-1} we get the matrix $U^{-1} = (u_{ij}^{\mp})$. As u_i are supposed to be unities, U^{-1} too is a matrix of the type (4.5).

$$E_{ik}^{\pm}(\lambda) = E_{ik}(\pm\lambda)$$

is an elementary matrix. Hence if we go from B by multiplying A from the left as well as from the right with a matrix of type (4.5) and (4.6) we conversely get A by multiplying B by matrices of the same type. B is said to be congruent to A , and is denoted by

$$A \sim B \quad (4.6)$$

Let $a \in A \cap B \neq 1$. It follows that $A \cap C = B \cap A = A \cap A = A$. Hence the matrix congruent to A (or a) as a coefficient of this congruence being congruent to any other (see Part II, § 2), is congruent to A (means to be identical) modulo m is independent of A .

[4.9] Let b_1, \dots, b_n be arbitrary but fixed elements of S and b_1, \dots, b_n be arbitrary elements of S . The elements

$$a_1 b_1 + \dots + a_n b_n \quad (4.9)$$

form a subalgebra of A , where $a_i \in A$ for $i = 1, \dots, n$. For any $c \in A$, the element c of S is congruent to any element of A modulo m , each element of A is congruent to c by the definition of m . But as it has been observed in § 1, the b_i are in A . Hence an element of S is congruent to c and any element of A is congruent to c . There exists therefore a correspondence between the subalgebra M of S and the set of all elements of A and every M is generated by the help of only one element.

We have therefore shown subalgebras generated by members of A . Let the determinants

$$\Delta_{1,1}, \dots, \Delta_{1,n} \quad (4.10)$$

be minors of degree k of the matrix A and let the elements of A be a_1, \dots, a_n of b_1, \dots, b_n . The minors $\Delta_{1,1}, \dots, \Delta_{1,n}$ belong to S , and for every fixed k of $1, \dots, n$ there is a subalgebra M_k . Let v_k be the highest common factor of the minors of M_k . In A , the minors of the elements of minority row a_1, \dots, a_n are minors of A of degree $k-1$. Hence

$$\Delta_{1,1} = a_1 \Delta_{1,1} + \dots + a_n \Delta_{1,1}$$

belongs to M_{k-1} . Hence in the sequence of minors

$$M_1, \dots, M_n$$

of minors we arrived at the present one and, therefore, by induction. Hence the Determinant $\Delta_{1,1} = a_1$ can be represented by the following manner

$$\delta_1 = a_1$$

$$\delta_2 = a_1 a_2$$

$$\delta_n = a_1 a_2 \dots a_n \quad (4.11)$$

$$\det A = \delta_n = a_1 a_2 \dots a_n$$

Theorem. Equivalent matrices have the same elementary divisors.

Proof. Any matrix obtained from the left side of (4.11) from the right side with a matrix U in such that the row i of U has 1 in position i in k , and with matrix V in such that the column j of V has 1 in position j in k , is obtained by unitary only and the matrix U and V are obtained by the row addition $A \rightarrow U^{-1} A U$, those u -ones of A in which the i -row is struck out are not so altered the same holds for the v -ones in which the j -th row as well as the i -th row occur. Let $A_{i,j}$ be a minor in which the i -row is not and the j -th row is not and let $A_{i,j+1}$ be a minor we get by expanding in $A_{i,j}$ the i -row and $A_{i,j+1}$ the j -th row then $A_{i,j+1}$ will be transformed into u in $U^{-1} A U$ and V into the row i of U .

$$E_{i,j}(\lambda) A \rightarrow E_{i,j}(-\lambda) (E_{i,j}(\lambda) A) = A.$$

M will be transformed into a submatrix M_1 of M , and M_2 of M into M_1 and M_2 , it follows that $M_1 = M_2$. Hence by a row addition M can be transformed, and the same holds accordingly for any column addition. Hence the theorem holds. Especially this holds for the minors of A in which i and j occur for every matrix of the same class.

By the help of matrices U and V in (4.11) and U the [4/4] following operation may be carried out:

(1) Interchange of two rows i and j in A is effected by

$$(a) \quad (a) \quad (a) - ((a) + (\beta)) = (-\beta) \quad (-\beta) \quad (b)$$

$$(b) \quad (a) + (\beta) \quad (a) + (\beta) \quad (a) \quad (a)$$

(2) Sweep out of i and j rows and columns by a series of operations (row additions if one of the rows is i say) and the j -th of them

$$a, ab_1, \dots, ab_n \rightarrow a, 0, \dots, 0$$

(3) If none of the elements of a row a_1, \dots, a_n is too large $1 \leq i \leq n$ then we can alter the row by column addition in such a manner that an element will appear with the property

$$N(b) \leq N(a_i) \text{ for every } i.$$

Proof. Let a_1 be an element for which $N(a_1) \leq N(a_i)$ for all i and a_2 not be divisible by a_1 , then it follows from (4.1) that there are elements a and b for which

$$a_2 + ca_1 = b, \quad N(b) \leq N(a_1) \leq N(a_i) \quad (4.12)$$

holds.



By our induction we can replace a_1 by b .

6. If a_1 is the least of the elements of the row and to the left of the elements of the column, then by our induction, we can change by our induction the element a_1 to b and then an element b will appear in the matrix for which $a_1 \leq b \leq a_1 + b$.

From (1) and (2) we see that we can suppose a_1 to be the first element of the first row and the first column and the first column to be swept out. As to the case where the least of the elements of the matrix will not be at (1, 1), there exists an element a_2 not divisible by a_1 at (1, 2) and a_2 is at (2, 1). By interchanging the rows and columns a_2 becomes the 1st element of the first row. Starting from this position we make row and column additions in the following manner:

$$\begin{pmatrix} a_1 & 0 \\ 0 & a_2 \end{pmatrix} \longrightarrow \begin{pmatrix} a_1 + a_2 & a_2 \\ 0 & a_2 \end{pmatrix} \longrightarrow \begin{pmatrix} a_1 & b \\ 0 & a_2 \end{pmatrix}.$$

With the help of these operations the "sweep out" can easily be done. Let a be an element of the matrix A for which that function N has a minimum value. If a is different from the least a_1 of the elements of A , then we can alter A in such a way that there appears an element b for which $a_1 \leq a_2 \leq b \leq a_1 + b$ and $b \neq a$. The procedure can be repeated. But the function N takes integral positive values only hence after a finite number of steps the procedure must stop, and that is only possible if an element of the matrix becomes equal to a_1 .

By (1) a_1 will be placed on the left upper corner of the matrix, and by the help of (2), the first row and the first column will be swept out. So we get

$$A \in (a_1, \dots)$$

The least element of A , a_1 , is at (1, 1) by (1), but the least b is equal to $a_1 + a_2$. By repeating the procedure we get therefore

$$A \in (a_1, \dots, a_2, \dots) \in (a_1, \dots, a_2, \dots, a_3, \dots)$$

The h, c, f of the minors of degree h is

$$h = c_1 \cdots c_{h-1} c_{h+1} \cdots c_n$$

$$\text{Hence, } c_1 \cdots c_n = c_1 \cdots c_{h-1} c_h c_{h+1} \cdots c_n \quad (4, 13)$$

$$c_h = c_1 \cdots c_{h-1} \text{ and}$$

$$A \in \left(\begin{array}{ccc} c_1 & & \\ & \ddots & \\ & & c_{h-1} \end{array} \right) \quad (4, 14)$$

$$c_h \text{ divisible by } c_{h+1} \cdots c_n \quad (4, 15)$$

If on the other hand (4, 14) is the determinant of some h rows of A have the value c_h by (4, 15) we get the following theorem.

Theorem. Every matrix A with elements from S except for a diagonal matrix (4, 14) where c_1 equals the product of all the elements c_i and if A is an elementary divisor of A then there is a correspondence between the rows of degree h and the sets of elementary divisors and every set of elements c_i satisfying (4, 16) is an admissible set.

From the preceding theorem it follows that if $A \in S$ and $A \in (4, 15)$ congruent to a diagonal matrix (4, 14) then use c_1, c_2, \dots, c_n and each c_i is a factor of a unity and therefore a unity (Def. 2). Hence A is a product of matrices of type (4, 6) and (4, 7) if A is determinate or unity. On the other hand the determinant of a product of matrices (4, 6) and (4, 7) is a unity. The matrix A whose determinate or unity is therefore identical with the products of matrices (4, 6) and (4, 7) and as to each matrix (4, 6) and (4, 7) there exists an inverse matrix of the same type. Every matrix whose elements belong to S and whose determinant is a unity of S has an inverse of the same type. Therefore the following theorem holds.

Theorem. $A \in U_2 B$ if and only if

$$A = C_1^{-1} B C_2 \quad (4, 16)$$

where C_1, C_2 are matrices with elements from S , their determinants being unities of S .

The transformation of a vector space by the matrix A is given by

$$y = Ax, \quad D \in \mathcal{D}_n(x)$$

Let

$$C_1(x) \equiv (y), \quad (x) = C_1^{-1}(y),$$

$$(x_1, \dots, x_n) = (y_1, \dots, y_n).$$

Then the normal form of the vector x is identical with the normal form of the vector y , the normal form of the vector x is identical with the normal form of the vector y . D is supposed to be a diagonal matrix with elementary divisors as elements and

$$(y) = D(x) \quad (4, 17)$$

holds

Proof. Let u_1, \dots, u_r be independent vectors in the Euclidean space, and k_1, \dots, k_r integral numbers. If the vectors $u_1 + k_1 u_2 + k_2 u_3 + \dots + k_{r-1} u_r$ are the only vectors for which $u_1 + k_1 u_2 + k_2 u_3 + \dots + k_{r-1} u_r = 0$ for the case that k is the ring of the integral numbers and u_1, \dots, u_r are the only vectors of the

[4/6] We shall now apply the theory of congruence matrices with elements from a ring S as it was introduced in § 1 to the theory of the congruence of matrices with elements from a field K . The properties of congruence invariance for transformation and the normal form given in § 2 will then appear in a new light.

Let S be a commutative ring and let A be a fixed matrix A and let C be a nonsingular matrix C . The polynomial $A - xE$ form a matrix $A - xE$ satisfying the properties (1), (2) and (3). The elements of A are the matrices of S . If C is a matrix with elements from A and $\det C \neq 0$, then

$$C^{-1}AC \in \mathcal{A}, \text{ and}$$

$$C^{-1}AC - xE = C^{-1}(A - xE)C \in \mathcal{A} - xE \quad (4, 18)$$

In other words, the elements of $\mathcal{A} - xE$ are congruent to $A - xE$ and therefore without any loss of generality we may assume that A has the normal form (2, 23) which consists of a diagonal system of submatrices A_1, \dots, A_r . Every A_i corresponds to a λ_i , the characteristic polynomial μ_i and it has the normal form (2, 24). This means that A_i consists of a diagonal system of submatrices each of them being of the type (2, 24), each element of the congruence A_i is equal to λ_i , the elements just above the diagonal are equal to 1 and all other elements are equal to 0. The reduction of $A - xE$ to a congruent matrix of the type (4, 16) will be effected by the help of the

following congruence:

$$\begin{pmatrix} a & 1 \\ 0 & a \end{pmatrix} \sim \begin{pmatrix} a' & 1-a'' \\ 0 & a' \end{pmatrix} \sim \begin{pmatrix} 1 & a'' \\ 0 & a' \end{pmatrix} \sim \begin{pmatrix} 1 & a'' \\ a' & a' \end{pmatrix} \sim \begin{pmatrix} 1 & a'' \\ 0 & a' \end{pmatrix} \quad (4.19)$$

and, for $(a, b) = au + bv = 1$,

$$\begin{pmatrix} a & 0 \\ 0 & b \end{pmatrix} \sim \begin{pmatrix} a & 0 \\ 0 & b \end{pmatrix} \sim \begin{pmatrix} 1 & 0 \\ 0 & 1 \end{pmatrix} \sim \begin{pmatrix} a' & 0 & 1 \\ 0 & b & 0 \end{pmatrix} \sim \begin{pmatrix} a' & 0 & 1 \\ 0 & b & 0 \end{pmatrix} \sim \begin{pmatrix} 1 & 0 \\ 0 & b \end{pmatrix} \quad (4.20)$$

We multiply λ_1 by 1 to $\lambda_1 + x$ and λ_2 by x to $\lambda_2 + x$ where x is put for $\lambda_1 - \lambda_2$. The number of occurrences of the submatrix x will be altered by these row and column additions. We start with the two first rows and

columns and change them to $\begin{pmatrix} 1 & 0 \\ 0 & (\lambda_1 - x) \end{pmatrix}$ the other elements being

unaltered. For x a function of this procedure, the different submatrices of A will correspond to congruent diagonal systems with the diagonal elements

$$\begin{aligned} 1, & \dots, 1, (\lambda_1 - x)^{e_1}, \\ 1, & \dots, 1, (\lambda_2 - x)^{e_2} \end{aligned} \quad (4.21)$$

$$1, \dots, 1, (\lambda_1 - x)^{e_1'} \quad (4.22)$$

where $e_1 \geq e_1', e_2 \geq e_2', e_1 + e_2 = e_1' + e_2'$

and λ_1 has the value $\lambda_1 + e_1 x$ and λ_2 has the value $\lambda_2 + e_2 x$ in the diagonal matrix with the elements (4.21), and a similar arrangement by interchanging the rows and the columns as follows:

$$1, \dots, 1, (\lambda_2 - x)^{e_2'}, \dots, (\lambda_1 - x)^{e_1'}, \dots, (\lambda_2 - x)^{e_2}, \quad (4.23)$$

From (4.22) it follows that (4.23) is the normal form of (4.1) if $\lambda_1 = \lambda_2$. As the alterations effected in the rows and columns of A do not alter the other submatrices of A we can reduce these submatrices independently to normal form (4.2). So we get a diagonal matrix congruent to A , the diagonal elements being powers of $\lambda - x$ with non-negative exponents. By interchanging the rows and columns and applying the congruence (4.20) we get a diagonal matrix

$$1, \dots, 1, \psi^k(x), \dots, \psi^l(x), \psi^l(x), \quad (4.24)$$

where $\psi = \sum_{i=1}^n \lambda_i f_i^{p-1}$ and $\lambda_i = \lambda_i^{p-1}$ is the same polynomial as f_i has been defined by (4.2). Every element $f \in A$ is divisible by the preceding elements. Hence (4.24) is the normal form $f = f_1 + f_2 A + \dots + f_n A^{n-1}$. If (4.24) is given we can find out the roots λ_i and the exponents $\alpha_i = p^{i-1}$ and by them the normal form (4.24) is uniquely defined. By these considerations we get the connection between the classes of matrices A formed with elements of a field K equivalent transformations and the classes of matrices formed with elements of the ring $K[x]$ and congruent to $A = xI$. There is an 1-1 correspondence between the classes. The normal form (4.24) of $A = xI$ is exactly given by the normal form (4.22) of A and conversely we get (4.22) from (4.24) by replacing the polynomials $\lambda_i = \lambda_i^{p^{i-1}}$ by products of powers of linear factors.

5. Here we consider a special case. Here we consider the following theorem

(5.1) Let K be a sub field of A and $[A : K] = 2$

Let σ be an automorphism of A not belonging to K (i.e. $K \neq A$). Then the fixed ring of σ is K and conversely in A we have $[A : K] = 2$ and σ is the only automorphism of A not belonging to K . If f is a polynomial in $K[x]$ of degree < 2 and $\sigma(f) = f$ then $f \in K$. Every automorphism (see Part II [2.2]) of A is a linear transformation which the elements of K are not altered. Transformations of f in $K[x]$ are of the form $f(x) \mapsto f(ax+b)$. Hence the transformations of A over K are of the form $f(x) \mapsto f(ax+b)$ with $a \neq 0$. In the first case every element of A is altered by the automorphism, hence the automorphism is the identity. In the second case every element of A is altered and σ is of the form $f(x) \mapsto f(ax+b)$ with $a \neq 0$. If f is a transformation

$$a \mapsto ba \longrightarrow a \mapsto ba \quad (5.1)$$

• an automorphism of A is the same the product of two elements
• transformations are the same the product of the corresponding elements
• $f(x) \mapsto f(ax+b)$ is an automorphism of A (where f is the identity) if and only if $a \neq 0$ and the elements of K are not altered. Hence the transformations of A are independent of the elements of A are depending on a and b are called a and b . The conjugate of f is denoted by

that

$$\bar{f} = f \quad (5.2)$$

conditions

$$\sum_i u_i^* u_j^* = 0 \quad \text{for } i \neq j \quad (5.5)$$

$$\sum_i u_i^* u_i^* = 1 \quad (5.6)$$

and
$$\lambda_i u_i^* = \varepsilon_i^* \quad \text{for } i = 1, \dots, n. \quad (5.7)$$

For $n = 1, \dots, n$, we multiply (5.5) by ε_i^* and $\sum_i \varepsilon_i^* \varepsilon_i^* = 0$. Then $\sum_i \varepsilon_i^* \varepsilon_i^* = \sum_i \varepsilon_i^* \varepsilon_i^*$ if (5.5) holds. If $n \geq 2$ there exists a non-trivial

vector $\varepsilon = (\varepsilon_1, \dots, \varepsilon_n)^T$ such that $\sum_i \varepsilon_i^* \varepsilon_i^* = 0$ and $\sum_i \varepsilon_i^* \varepsilon_i^* = 1$ holds, for $i = 1, 2, \dots, n$, $i \neq j$. We can assume that $\varepsilon_i^* \varepsilon_i^* = 1$ for $i = 1, 2, \dots, n$. We get $\varepsilon_i^* \varepsilon_i^* = \varepsilon_i^* \varepsilon_i^*$ and $\varepsilon_i^* \varepsilon_i^* = 1$ for $i = 1, 2, \dots, n$.

$$(\varepsilon_1^*, \dots, \varepsilon_n^*) \neq (0, \dots, 0), \quad \sum_i \varepsilon_i^* \varepsilon_i^* = 0, \quad \text{for } i \neq j$$

$\sum_i \varepsilon_i^* \varepsilon_i^* = 1$ if and only if $n = 1$. If $n \geq 2$, $\varepsilon_i^* \varepsilon_i^* = 1$ satisfies the conditions (5.5), (5.6) and (5.7).

For $n = 1, 2, \dots, n$, we assume that $\varepsilon_i^* \varepsilon_i^* = 1$ for $i = 1, 2, \dots, n$ and $\varepsilon_i^* \varepsilon_i^* = 0$ for $i \neq j$ and $n = 2$ and (5.7) holds.

[5/8] The conditions (5.5) and (5.7) can be considered as a property of the matrix $\varepsilon_i^* \varepsilon_i^*$. To express properties of this kind in a simple form it is helpful to use the following notations:

Let $\Lambda = (\varepsilon_i^* \varepsilon_j^*)$ be a matrix with elements from Λ then

$$\Lambda = (\varepsilon_i^* \varepsilon_j^*) \quad \text{for } i, j = 1, 2, \dots, n, \quad \text{is the transpose of } \Lambda, \quad (5.8)$$

$$\Lambda^* = (\varepsilon_i^* \varepsilon_j^*)^* \quad \text{is the conjugate of } \Lambda \quad (5.9)$$

$$\Lambda^\dagger = (\Lambda^*)^* = \Lambda \quad (5.10)$$

From these formulas it follows that

$$(\Lambda B)^* = B^* \Lambda^*, \quad \Lambda B = A B, \quad (\Lambda^\dagger)^* = B^\dagger \Lambda^\dagger$$

and for $\det \Lambda \neq 0$,

$$\Lambda^{-1} = (\Lambda^*)^{-1}, \quad \Lambda^{-1} = \Lambda^{*-1}, \quad (\Lambda^\dagger)^{-1} = \Lambda^{-1\dagger} \quad (5.11)$$

The equations (5.11), (5.12), and (5.13) are therefore equivalent with

$$((u_i^j))^\dagger = ((u_i^j))^{-1}. \quad (5.12)$$

A matrix satisfying these conditions is said to be *unitary*. As the transpose of an (nonsingular) matrix is also unitary, a unitary matrix $((u_i^j))$ also satisfies

$$\sum_j u_i^j \bar{u}_j^k = 0, \quad \text{for } i \neq k, \quad (5.14)$$

$$\sum_j u_i^j \bar{u}_i^j = 1. \quad (5.15)$$

As $\det ((u_i^j)) = \det ((u_i^j))^\dagger = \det ((u_i^j)^{-1}) = 1/\det ((u_i^j))$ holds, hence

$$\det ((u_i^j)) = \pm 1. \quad (5.16)$$

A unitary matrix remains unitary after an arbitrary permutation of the rows and/or of the columns. The product of unitary matrices is unitary.

$$\text{If } U = \begin{pmatrix} \pm 1 & \\ & \overline{U'} \end{pmatrix} \quad (5.17)$$

and one of the matrices U, U' is unitary, the other is unitary, too. If the elements of a unitary matrix are elements of \mathbb{R} then $U = U'$ and the matrix is said to be *real unitary*, and if the elements are complex, the relation used in (5.17) is (5.18), the term of *unitary matrices* is preferred to the theorem (5.16) in the following section.

Theorem. To every vector space V there is a unique matrix A with the property that $u \cdot g$ is the row u multiplied by A for $g \in V$ if and only.

If the element \bar{a} of a module M and the element a of M are $\bar{a} = \bar{a}$ and $a = a$ polynomial* belong to A and

$$H = H^\dagger \quad (5.19)$$

H is said to be an *Hermitian matrix*.

Let H be an Hermitian matrix, U a unitary matrix, then $H_1 = U^{-1} H U$ has the same characteristic polynomial as H and $H_1^\dagger = (U^{-1} H U)^\dagger = U^\dagger H^\dagger U = H_1$. Hence H_1 is Hermitian. Let λ be an arbitrary root of $\chi_H(x)$ and $\lambda \neq 0$.

* This condition is equivalent to the condition that A is a closed field, e.g. \mathbb{C} or a field of the complex numbers.

a vector for which $(H - \lambda I)x = 0 \neq 0$ holds. Such a vector (β) must exist since $2 - 1 = 1$. Let U be a unitary matrix whose first column differs from (β) only by a factor $\neq 0$. By $H_1 = U^{-1} H U$ the first unit vector is transformed in the same manner as β is transformed by H , hence it is multiplied only by λ . The first column of H_1 has therefore the elements $\lambda, 0, \dots, 0$ and as H_1 is Hermitian,

$$H_1 = \begin{pmatrix} \lambda & \\ & H \end{pmatrix}$$

Hence λ is an element of K . H is an Hermitian matrix of degree $n - 1$. We can transform it by a unitary matrix in the same manner as H has been transformed

$$U_1 H_1 U_1^{-1} = \begin{pmatrix} \lambda' & \\ & H' \end{pmatrix} \quad \text{As the matrix } U_1 = \begin{pmatrix} 1 & \\ & I \end{pmatrix} \text{ is}$$

$$\text{also unitary} \quad U_1 H_1 U_1^{-1} = \begin{pmatrix} \lambda' & \\ & H' \end{pmatrix} \quad \text{is an Hermitian matrix}$$

Afterwards we get H transformed into a diagonal matrix by a matrix which is the product of unitary matrices and is therefore unitary. So we get the following important theorem

Theorem. An Hermitian matrix H can be transformed by a unitary matrix into a diagonal matrix and the roots of $\chi_H(x)$ belong to K .

Let K be the field of the real numbers, A be the field of the complex numbers, then it follows from this theorem that the roots of the characteristic polynomial of a matrix are a symmetric real matrix. Hence we can consider an matrix with real elements for which $a_{ij} = a_{ji}$ holds as Hermitian, where $K = A$ is equal to the field of the real numbers. Hence the roots of the characteristic polynomial belong to A . Hence we can apply the preceding theorem on this case, the unitary matrices becoming now orthogonal matrices and we get the

Corollary. If in a matrix $x \in A$ with real numbers as elements, $a_{ij} = a_{ji}$ holds, then $\chi_x(x)$ has real roots only and A can be transformed by an orthogonal matrix with real coefficients into a diagonal matrix.

The theory of matrices will now be applied to linear and quadratic [6/4] forms. We introduce a finite set of indeterminates

$$\begin{aligned} x_1, x_2, \dots, x_n & \quad y_1, y_2, \dots, y_n \\ \bar{x}_1, \bar{x}_2, \dots, \bar{x}_n & \quad \bar{y}_1, \bar{y}_2, \dots, \bar{y}_n \end{aligned} \quad (5, 16)$$

Corresponding elements of the two lines will be said to be conjugate. They are supposed to be different if $[A:K] = 2$, and to be identical if $K = A$. The automorphism of A defined in [5.1] by which every element of A is replaced by its conjugate will now be extended to an automorphism of the ring

$$\Sigma = A[x_1, \dots, x_n, \bar{x}_1, \dots, \bar{x}_n] \quad (5, 17)$$

of the polynomials in the indeterminates (5.16) with coefficients from A . This extension will be made so that every indeterminate (5.16) hereafter be replaced by its conjugate. So we have simply to replace in every polynomial each coefficient and each indeterminate by its conjugate. In the case $K = A$ the automorphism is the identity. In every case there belongs to every element x of Σ a uniquely defined conjugate element \bar{x} and $x = \bar{\bar{x}}$ holds.

In [1, §2] a vector has been represented (see [1.25]) by a matrix, the first column of which is formed by the coordinates, the elements of the other columns being 0. We will now apply the notations of conjugate matrix and transposed matrix to these special matrices, so that

$$(x) = \begin{pmatrix} x_1 & 0 & 0 \\ \vdots & & \\ x_n & 0 & 0 \end{pmatrix}, \quad (\bar{x}) = \begin{pmatrix} \bar{x}_1 & 0 & 0 \\ & & \\ & 0 & 0 \end{pmatrix} \quad (5, 18)$$

$$(x)^{\cdot} = \begin{pmatrix} \bar{x}_1 & \bar{x}_2 \\ 0 & 0 \\ 0 & 0 \end{pmatrix}, \quad (\bar{x})^{\cdot} = \begin{pmatrix} x_1 & x_2 \\ 0 & 0 \\ 0 & 0 \end{pmatrix}$$

Let $A = (f_{ij})_1^1$, then $A(y)$ is a matrix with the first element

$$\sum a_{ij}^1 x_i y_j \quad (5, 19)$$

all other elements being equal to 0. To every matrix A there corresponds

a bilinear form (19) and conversely let

$$(x) = B(x'), \quad y = C(y'), \quad \text{and}$$

$$B^*AC = A' = (a'_{ij}), \quad (5, 20)$$

then $(x')^* B^* AC(y') = (x')^* A(y).$

Hence $\sum a_{ij} x_i y_j = \sum a'_{ij} x'_i y'_j. \quad (5, 20')$

The formulae (5, 20) and (5, 20') give the transformation of a bilinear form.

We will now consider the case where A is an Hermitian matrix, and where x and y are conjugate; then B and C are also conjugate, (5, 19) becomes an Hermitian form

$$\sum a_{ij} x_i \bar{x}_j, \quad \text{where} \quad a_{ij} = \bar{a}_{ji}, \quad (5, 21)$$

and $\sum a_{ij} x_i x_j = \sum a'_{ij} x'_i x'_j$, where $a'_{ij} = A = C + AC \quad (5, 22)$

A is therefore an Hermitian matrix too. In the application we wrote $K = A$, $a_{ij} = a'_{ij}$ and $x_i = x'_i$, the bilinear form becomes quadratic and the Hermitian matrix becomes a symmetric one.

$$\sum a_{ij} x_i x_j = \sum a_{ij} x_i^2 + 2 \sum_{i < j} a_{ij} x_i x_j. \quad (5, 23)$$

If the characteristic of K is different from 2 every quadratic form in x_1, \dots, x_n can be represented in the form (5, 23). We will omit the case of characteristic 2 which needs a special treatment; hence there is an 1:1 correspondence between the quadratic forms and the symmetric matrices. The transformation is done by

$$a'_{ij} = A = C^*AC, \quad x_i = C x'_i, \quad \sum a_{ij} x_i x_j = \sum a'_{ij} x'_i x'_j, \quad (5, 24)$$

This formula for the transformation of quadratic forms reminds us of the transformation of matrices [see (3, 24)] and the notion of inverse matrix has been replaced here by the notion of transposed matrix and it is not necessary that $\det C \neq 0$. On applying the theorem of (5, 3) and its corollary to (4, 22) and (5, 24) we get the following fundamental theorem

Theorem. Every Hermitian form can be transformed by a unitary transformation into the normal form

$$\sum \alpha_i x_i \bar{x}_i. \quad (5, 25)$$

and every quadratic form with coefficients from K can be transformed by an orthogonal transformation with coefficients from K into the normal form

$$\sum a_i x_i^2 \quad (5, 26)$$

a_i being elements of K not with the cases

Let $K = \mathbb{R}$ be the field of the real numbers, then every quadratic form (5, 5) can be transformed into

$$Q(x) = a_1 x_1^2 + \dots + a_r x_r^2, \quad (5, 27)$$

where the a_i are non-zero different from zero by the transformation with an orthogonal matrix A with determinant ± 1 an orthogonal transformation is equal to ± 1 . The rank of the matrix A will not be altered by the transformation and is therefore equal to r . If we replace x by $-x$ the formula (5, 27) will not be altered, but the sign of the orthogonal transformation becomes -1 . Hence we can transform an arbitrary quadratic form into a normal form (5, 27) by an orthogonal transformation with determinant ± 1 . Since a coordinate change means in the geometry of the space of n dimensions a rotation through the origin. This transformation is well known in the analytic geometry — comes and is called quadratic as the transformation to the principal axes.

Let $a_i = +b_i^2$, and $b_i x_i = y_i$, then (5, 27) is transformed to a sum of squares with certain signs \pm . After a permutation of the indices and replacing y_i by x_i we get therefore

$$q(x) = x_1^2 + \dots + x_r^2 - x_{r+1}^2 - \dots - x_s^2. \quad (5, 28)$$

Hence every quadratic form can be transformed to (5, 28) by a non-degenerated linear transformation with real coefficients. The integer r is the rank of matrix of the quadratic form and therefore invariant. We will prove that p is an invariant too.

Theorem. Every quadratic form with real coefficients can be transformed into the normal form (5, 28) by a non-degenerated linear transformation with real coefficients.

Proof. As it has been shown above the transformation into the normal form is always possible and r is an invariant. We have therefore to prove that a transformation of $q(x)$ into

$$q(x) = x_1^2 + \dots + x_r^2 - x_{r+1}^2 - \dots - x_s^2$$

by linear non-degenerated substitution is possible only if $p=q$. Let $p \neq q$, say $p > q$ without any loss of generality.

$$x_i = b_1^i x_1 + \dots + b_r^i x_r,$$

$$x_i = c_1^i x_1 + \dots + c_r^i x_r, \quad i=1, \dots, r$$

$q(x) = q_1(x)$ for corresponding systems (x_1, \dots, x_r) and (x_1, \dots, x_r) . For $q+r-p < r$ linear homogeneous equations

$$c_1^k x_1 + \dots + c_r^k x_r = 0 \quad k=1, \dots, q$$

$$x_i = 0 \quad i = p+1, \dots, r$$

have a solution $(\xi_1, \dots, \xi_r, 0, \dots, 0)$ different from $(0, \dots, 0)$ viz. the rank of the matrix of this system of equations is $\leq q+r-p < r$. The corresponding values of x_1, \dots, x_r are $(\xi_1, \dots, 0, \xi_{p+1}, \dots, \xi_r)$ and different from $(0, \dots, 0)$. Hence $\xi_1 \xi_2 \dots \xi_p > 0$, $q_1(\xi) < 0$ in contradiction to $q(\xi) = q_1(\xi)$.

A quadratic form is said to be *positive definite* if $n=r=p$, it is *negative definite* if $n=r$, $p=0$, it is *semidefinite* if $n > r$, $p=r$ or 0 , and it is *indefinite* if $r > p > 0$.

[5/11]

Finally we will give a geometrical interpretation of the last results without going into the details.

In the projective $(n-1)$ -dimensional space a quadric becomes represented by

$$\rho \sum a_i^2 x_i x_i = 0,$$

where $\rho \neq 0$ is an arbitrary factor.

Hence the quadric has one and only one normal form

$$\rho q_1(x) = 0,$$

and the sign of ρ can be fixed in such a manner that $q(x)$ has not fewer positive than negative terms. We get therefore the different types of quadrics in the projective $(n-1)$ -dimensional space given by the different normal-forms

$$q(x) = 0 \quad \text{for } r=1, \quad n-1/2 \leq p \leq r \quad (5, 20)$$

Especially the quadrics without any real point are those for which $p=r=n$. The normal form (5, 20) has the property that every fundamental point of the coordinate system (i.e. every point, for which all the coordinates are equal to 0 except x_i), is polar to the opposite hyperplane $x_i = 0$.

In the affine n -dimensional space the quadrics are given by

$$\sum a_i x_i x_i + \sum b_i x_i + c = 0.$$

On applying the theorem of [5] we get easily the following types of affine normal forms.

$$\begin{aligned} q(x) &= 0, & r &= 1, \dots, n, & r/2 \leq p \leq r \\ q(x) &= 1, & r &= 1, \dots, n, & r \leq r \\ q(x) &= a_{r+1}, & r &= 1, \dots, n-1, & r/2 \leq p \leq n \end{aligned} \quad (5, 20)$$

If we replace $q(x)$ by the quadratic form $Q(x)$ of (5, 27) we get the types of quadrics different in the sense of metric geometry. The formula (5, 27) can also be interpreted for projective geometry.

Let $R(x_1, \dots, x_n)$ and $S(x_1, \dots, x_n)$ be two quadratic forms. S representing a quadric without real points, then we transform both coordinates in such a manner that S is transformed to $S = x_1^2 + \dots + x_n^2$, and R to R' . By any orthogonal transformation S will not be altered, but we can transform R by an orthogonal transformation to the normal form (5, 27). Hence we can transform simultaneously

$$\begin{aligned} R &\text{ to } a_1 x_1^2 + \dots + a_n x_n^2, \\ S &\text{ to } x_1^2 + \dots + x_n^2 & \text{and the pencil} \\ \lambda R + \lambda S &\text{ to } (\lambda a_1 + \lambda) x_1^2 + \dots + (\lambda a_n + \lambda) x_n^2 \end{aligned}$$

The elements of the pencil have therefore been transformed to the normal-form simultaneously.

§ 6. RESULTANTS

Let

{6/1}

$$f(x) = x^n + a_1 x^{n-1} + \dots + a_n = (x - \alpha_1) \dots (x - \alpha_n) \quad (6, 1)$$

$$g(x) = x^m + b_1 x^{m-1} + \dots + b_m = (x - \beta_1) \dots (x - \beta_m) \quad (6, 2)$$

$$R(f, g) = \prod_i \prod_j (\alpha_i - \beta_j). \quad (6, 3)$$

Then $R(f, g)$ is uniquely defined by the polynomials f and g and it is said to be the resultant of f and g . The necessary and sufficient condition that f and g may have a common root is that the resultant is equal to zero.

From (6, 3) it follows

$$R(f, g) = (-1)^{mn} R(g, f); \quad (6, 4)$$

from (6, 2) and (6, 3)

$$R(f, g) = \prod_{i=1}^n g(\alpha_i), \quad (6, 5)$$

and by interchanging f and g and applying (6, 4) we get

$$R(f, g) = (-1)^{mn} \prod_{i=1}^m f(\beta_i). \quad (6, 6)$$

The right side of the equation (6, 5) is a symmetric polynomial in $\alpha_1, \dots, \alpha_n$ with $\alpha_i = a_i + b_i x$, $b_i = b_1 + \dots + b_m$ these polynomials b_i having integral coefficients. From Part II [1, 3] it follows that $R(f, g)$ can be represented as a polynomial in the elementary symmetric polynomials of $\alpha_1, \dots, \alpha_n$ with coefficients $\varphi_i = \varphi_i(b_1, \dots, b_m)$ where φ_i has integral coefficients. As the elementary symmetric polynomials of $\alpha_1, \dots, \alpha_n$ are equal to f_1, \dots, f_n , the resultant $R(f, g)$ can be represented as a polynomial in $a_1, \dots, a_n, b_1, \dots, b_m$ with integral coefficients.

For a reason which will become evident later on this representation will be written as

$$R(f, g) = R(a_1, \dots, a_n; b_1, \dots, b_m). \quad (6, 7)$$

If in any term $A = a_1^{t_1} a_2^{t_2} \dots a_n^{t_n} b_1^{t_{n+1}} b_2^{t_{n+2}} \dots b_m^{t_{n+m}}$ the factors a_i are represented by α_i and the factors b_i by β_i , then A becomes an homogeneous polynomial in α_i and β_i of degree *

$$t = t_1 + 2t_2 + \dots + nt_n + t_{n+1} + 2t_{n+2} + \dots + mt_{n+m} \quad (6, 8)$$

It is said to be the weight of A . From (6, 7) it follows that $R(f, g)$ is homogeneous of degree mn . Hence each term of $R(a_1, \dots, a_n; b_1, \dots, b_m)$ has the weight mn . From (6, 8) we see that one of these terms is equal to b_m^{mn} .

Let S be a polynomial in $a_1, \dots, a_n, b_1, \dots, b_m$ with the property that S_1, \dots, S_{n+m} are all zero if $a_1, \dots, a_n, b_1, \dots, b_m$ have a common root. By representing a_i as symmetric polynomials f_1, \dots, f_n and b_i as symmetric polynomials

* See Part II, p. 19

in β_1, \dots, β_m we get

$$S = \Sigma = \Sigma (a_1, \dots, a_n, \beta_1, \dots, \beta_m)$$

The right side of this equation is equal to x^i if $a_1 = \beta_1$

Hence

$$\Sigma = \Sigma (a_1, \dots, a_n, \beta_1, \dots, \beta_m) = \Sigma (x_1, \dots, x_n, a_1, \dots, \beta_m)$$

Subtracting the corresponding terms on the right side we see that Σ is divisible by $a_1 - \beta_1$ and in the same manner it follows that Σ is divisible by $a_i - \beta_i$, hence Σ is divisible by $R(f, g)$. Or $1/R(f, g) = \Sigma/R(f, g)$. From the 2nd theorem of Part II, 10th, it follows therefore that $R(1, a_1, \dots, a_n; 1, b_1, \dots, b_m) = (S, R(1, a_1, \dots, a_n, 1, b_1, \dots, b_m))$. Hence S is divisible by the resultant. The weight of every term is therefore not less than $m + n$, if each term has the weight $m + n$ differs from the resultant by a factor of weight 0 only and this factor is the coefficient of the term b_m^n in S .

To get the resultant as a polynomial in $a_1, \dots, a_n, b_1, \dots, b_m$ we have [6/2] therefore to find out a polynomial S , with the following three properties

1. $S \neq 0$ if f and g have a common root
2. Each term of S has the weight $m + n$
3. The term b_m^n has the coefficient 1

A polynomial of this kind can easily be found out by the following consideration

Let $f(x) = 0 = g(x)$ then the following $n + m$ equations hold

$$a_0 f(x) = a_0^{n+1} + a_1 a_0^{n+1-1} + \dots + a_n a_0^n = 0$$

$$a_0^{n+1} f'(x) = a_1 a_0^{n+1-1} + \dots + a_{n-1} a_0^{n+1-n} + a_n a_0^{n+1-n-1} = 0$$

$$\dots$$

$$a_0^n f^{(n)}(x) = a_n a_0^{n+1-n} + \dots + a_{n-1} a_0^{n+1-n-1} + a_n a_0^{n+1-n-1} = 0$$

$$a_0^n g(x) = a_0^{m+1} + b_1 a_0^{m+1-1} + \dots + b_m a_0^m = 0$$

$$a_0^{m+1} g'(x) = a_1 a_0^{m+1-1} + \dots + b_{m-1} a_0^{m+1-m} + b_m a_0^{m+1-m-1} = 0$$

$$\dots$$

$$a_0^m g^{(m)}(x) = a_m a_0^{m+1-m} + b_1 a_0^{m+1-m-1} + \dots + b_m a_0^{m+1-m-1} = 0$$

We consider this system as a system of n linear equations in a_1, \dots, a_n . It can be satisfied only if its determinant is equal to zero. Hence a necessary condition for that f and g may have a common root is

$$S = \begin{vmatrix} 1 & a_1 & \dots & a_n \\ & 1 & \dots & a_{n-1} \\ & & \ddots & \vdots \\ & & & 1 & a_1 & \dots & a_n \\ & & & & 1 & \dots & a_{n-1} \\ & & & & & \ddots & \vdots \\ & & & & & & 1 & a_1 & \dots & a_n \\ & & & & & & & 1 & \dots & a_{n-1} \\ & & & & & & & & \ddots & \vdots \\ & & & & & & & & & 1 & a_1 & \dots & a_n \end{vmatrix} \quad (6, 9)$$

$$= 0$$

The terms of S are of weight n each; the term b_n^n is the diagonal element and has the coefficient 1. Hence

$$S = R(f, g). \quad (6, 10)$$

[6, 11]

Let $F(x) = a_n x^n + \dots + a_0$

$G(x) = b_n x^n + \dots + b_0$

where nothing is supposed about the coefficients. We define now

$$R(F, G) = R(a_n, \dots, a_0, b_n, \dots, b_0)$$

$$a_n \dots a_0$$

$$= \begin{vmatrix} 1 & a_n & \dots & a_0 \\ & 1 & \dots & a_{n-1} \\ & & \ddots & \vdots \\ & & & 1 & a_n & \dots & a_0 \\ & & & & 1 & \dots & a_{n-1} \\ & & & & & \ddots & \vdots \\ & & & & & & 1 & a_n & \dots & a_0 \\ & & & & & & & 1 & \dots & a_{n-1} \\ & & & & & & & & \ddots & \vdots \\ & & & & & & & & & 1 & a_n & \dots & a_0 \end{vmatrix} \quad (6, 10)$$

$$b_n \dots b_0$$

$$b_n \dots b_0$$

As in (6, 10) there are m rows with elements a_i and n rows with elements b_i . If $a_n = b_n = 1$, then $F = f$, $G = g$, and we see from (6, 9) and (6, 10) that the notation $R(F, G)$ conforms to $R(f, g)$. We have to consider three cases

1. $a_n \neq 0, b_n \neq 0$.

$$F(x) = a_n \left(x^n + \dots + \frac{a_0}{a_n} \right) = a_n \phi(x) = a_n (x - \alpha_1) \dots (x - \alpha_n)$$

$$G(x) = b_n \left(x^n + \dots + \frac{b_0}{b_n} \right) = b_n \psi(x) = b_n (x - \beta_1) \dots (x - \beta_m)$$

From (8,10) it follows that

$$R(F, G) = a^m b^n R(\phi, \psi). \quad (8, 11)$$

Hence for (6, 3), (6, 5), and (6, 6)

$$R(F, G) = a^m b^n \prod_{i=1}^m (a_i - a_i') \prod_{j=1}^n (b_j - b_j') = (-1)^{m+n} b^n \prod_{j=1}^n (\beta_j) \quad (8, 12)$$

Hence in this case $R(F, G) = 0$ is the necessary and sufficient condition for the existence of a common root of F and G .

2. $a_m = 0, b_n = 0$. From (8,10) it follows that $R(F, G) = 0$, independent of the existence of a common root.

3. $a_m \neq 0, b_n = 0$ (or $a_m = 0, b_n \neq 0$).

Let $b_1 = \dots = b_{n-1} = 0$, then $R(F, G) = 0$ and every root α' of F satisfies obviously $G(\alpha') = 0$.

Let every $b_{i+j} = 0, b_n \neq 0, 1 \leq i \leq m, j \geq 1$ and $b_{i+j} = b_{i+j-1} + b_n$ for $i = 1, \dots, m$.

By setting 0 for b_n in (8,10) we get $R(F, G) = a_m R(F, G_1)$

and by mathematical induction $R(F, G) = a_m R(F, G)$

Hence $R(F, G) = 0$ if and only if F and G_1 have a common root. Since F and G have a common root, the corresponding expansion holds for $i = 0, b_n \neq 0$. By these considerations we get the following theorem:

Theorem. If F and G have a common root, then $R(F, G) = 0$. If $R(F, G) = 0$, then either $a = b = 0$ or F and G have a common root.

Example 1. Consider the case $m = 1, n = 1$, and (8,12) in the case 3.

$$1. \text{ Let } F(x, y) = \sum_{i=0}^m a_i x^i y^0 \text{ and } G(x, y) = \sum_{j=0}^n b_j x^0 y^j.$$

State the necessary and sufficient condition for $F(x, y) \neq 0$.

Let $a_1, \dots, a_m, b_1, \dots, b_n$ be the factors of the last column of the $(m+1) \times (n+1)$ determinant (8,10), and let

$$U(x) = 0, x^{m-1} + \dots + u_{m-1}x + u_m, \quad V(x) = 0, x^{n-1} + \dots + v_{n-1}x + v_n.$$

* See Part I, p. 22



CORRECTIONS

Part I. (see the corrections given in Part II.)

Page.	Line.	Read	For
iii (Preface)	19	does	do
13	19	(2/H)	(2)
16	14	depend on	apply to
20	10	10.	(11)
	23	11.	10.

Part II.

6	10	$c + 0$	$c + 0$
6	19	M	M'
11	22	$(b' + w)$	$(b' + t)$
	23	$a'w$	$a't$
13	20	are	and also the distributive law are
	22	a non-distributive system	a system
	23	the reader may verify that (2, 8)	we should only
		generate an addition and a	prove that
		multiplication of the classes for	
		which the commutative, associative, and distributive laws	
		hold, and	
18	21	field	field
18	11	a_{n-1}	a_{n+1}
21	12 and 13	interchange the exponents :	
23	24	$f(x), \phi(x), \psi(x)$	$f(x), \phi(x), \psi(x)$
26	10	$\frac{a_n}{b_n}$	$\frac{b_n}{a_n}$
	11	$\frac{a_{n+1}}{b_n}$	$\frac{b_n}{a_{n+1}}$
	26	$\phi(x) \phi_1(x)$	$\phi(x) \psi_1(x)$
33	6	K	K_1'
	26	$K(\sigma)$	$K(a)$
38	9	N	K [three times !]

Page.	Line.	Read	For.
42	18	to $K(a)$	to K
46	9	a_i	a_i
47	19	$x^{n-1} + \dots + b_n$	$x^{n-1} + \dots + b_n$
	30	$F(a_1, \dots, a_n)$	$F(a_1, \dots, a_n)$
52	17	h	h [error not in all copies]
61	8	(13, 2)	(13, 2)
	5	(13, 2)	(13, 1)
	6	$n(n-1)$	$n(n-1)$

Parts III--V. (in this volume)

3	19	P'	P [twice]
	20	Q'	Q [twice]
	25	a_{i+i-1}	a_{i+i-1}
8	25	$a-x$	$a+1-a$
11	5	$>a>$	$<a<$
12	4	$+$	$-$ [between the fractions]
13	27	$\frac{P_{2n}}{Q_{2n}}$	$\frac{P_{2n}}{P_{2n}}$
14	17	$s_n \lambda$	s_n, λ
16	1	a	a
	15, 19	purely periodic	purely
19	8	[8/4] [on the margin]	
	9	>1	>0
20	12	$\alpha+\beta$	$\alpha_1+\beta_1$
		$\alpha+\alpha$	$\alpha+\alpha_1$ [twice]
22	5	$\sqrt{26}$	26
31	1	Σa_i	s_i
	6	If this sum is convergent, Hence the sum taken for odd indices is divergent, where it follows easily that $Q_{2n+1} \rightarrow \infty$. Hence	
	10	interchange "then" with "and"	

Page.	Line.	Read	For
41	17	(1, 3)	(3)
42	21	$(x-a)(x-\bar{a})$	$(x-a)(x-\bar{a})$
53	16	(1, 30)	(30)
58	5	2-40824	2-40224
59	15	$\left(1 + \frac{2b_2^m + b_4^m}{b_1^m}\right)$	$\frac{(1 + 2b_2^m + b_4^m)}{b_1^m}$
61	14	b_1	b_2
66	15	$[5/2]$	$[6/]$
68	22	Δ_{2-2}	$\Delta_{2,1}$
72	7	(see Part II [1/2])	(see Part II [1/2],
	18	modul M	modul m
	30	$\{(e_i^j)\}$, where $e_i^i = 1$, and e_i^j	$\{e_i^j\}$, where $e_i^i = 1$, and e_i^j
74	7	exists	exist
75	18	a_i^j	a_i^j
76	21	exists	exist
79	26	[omit=]	
80	10	[place λ from the 2nd to the 1st column in the 1st line of (2, 14)]	
	11	$\Delta^{(r)}$	$\Delta^{(n)}$
		$\chi_{A^{(r)}}$	$\chi_{A^{(n)}}$
82	14	[replace the index m by q]	
85	9	$W_{1,1}$	$W_{1,1,1}$
101	17	exists	exist
105	23	$(x)^*A(y)$	$A(y)$